

AFOSR 70 -28 09 TR

AD716591

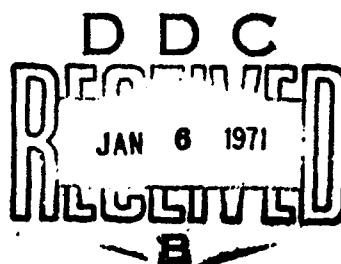
STATISTICS LABORATORY

MATHEMATICS DEPARTMENT

THE OHIO STATE UNIVERSITY

COLUMBUS, OHIO

L This document has been approved for public
release and sale; its distribution is unlimited.



Reproduced by
NATIONAL TECHNICAL
INFORMATION SERVICE
Springfield, Va. 22151

PRINCIPAL COMPONENT ANALYSIS
OF TIME SERIES^{1,2}

J. Richard Stewart
Division of Statistics
The Ohio State University

September, 1970

1. Presented in partial fulfillment of the requirement for the Degree of Philosophy in the Graduate School of The Ohio State University.
2. Research partially sponsored by the Air Force Office of the Scientific Research, Office of Aerospace Research, United States Air Force, under AFOSR Grant No. AF-AFOSR-1305-67. The United States Government is authorized to reproduce and distribute reprints for Governmental purposes not withstanding copyright notation hereon.

ACKNOWLEDGMENTS

I wish to thank Professors D. Ransom Whitney and Jagdish Kustagi for their help toward the completion of this dissertation, and especially Dr. Peter O. Anderson for his assistance throughout its writing. I also wish to thank my wife, Sarah, for typing the text and for her help with the numerical computations.

INTRODUCTION

A situation which sometimes arises is that in which a time series $y(t)$, $t = 0, 1, \dots$ is to be modelled as a function of another time series $x(t)$, $t = 0, 1, \dots$ (or more generally as a function of k series $x_1(t), \dots, x_k(t)$; $t = 0, 1, \dots$). It is often the case that for any time t , $y(t)$ will depend not only on $x(t)$, but also on $x(t - 1)$, $x(t - 2)$, \dots , $x(t - r)$ for some integer $r \geq 0$, and sometimes on the complete past history of $x(t)$. When all of the series are second order stationary and a large number of observations is available, the powerful methods of spectral analysis may be employed to estimate various types of linear models. On the other hand, when relatively few observations are available and the series involved may be nonstationary, at least to the extent of having trending means, these methods may not be appropriate, and yet it may be desirable to have some sort of frequency decomposition of the series. (For instance, it may be of interest to model trends and stationary parts of the series separately.)

The purpose of this study is to investigate some of the properties of a method whereby $y(t)$ is modelled not directly as a linear function of $x(t)$, $x(t - 1), \dots, x(t - r)$, but rather as a linear function of some of the principal components of $x(t), x(t - 1), \dots, x(t - r)$. Chapter I, formulated in terms of a set of random variables y, x_1, \dots, x_p , investigates the covariance structure among these variables which will yield the result that the first $m < p$ principal components of x_1, \dots, x_p provide a better predictor of y than does any subset x_{i_1}, \dots, x_{i_m} .

In chapter II, principal component processes of $x(t), x(t - 1), \dots, x(t - n)$ are defined and their properties noted. It is shown that when $x(t)$ is one of the usual types of second order stationary processes, the principal component processes effect a partial frequency decomposition of $x(t)$.

In chapter III, it is shown that for certain processes containing deterministic components, certain of the principal component processes tend to filter out these deterministic components.

An example of the behavior of the principal component processes of an artificially generated process is shown in the appendix together with the principal components of an autoregressive process, some of the transfer functions corresponding to the eigenvectors of a moving average process, and the first two principal component processes derived from a series representing quarterly total U.S. personal income.

CHAPTER I

REGRESSION ON PRINCIPAL COMPONENTS

Let X be the p component random vector

$$\begin{bmatrix} x_1 \\ \vdots \\ x_p \end{bmatrix}$$

with covariance matrix Σ . It will be assumed that $EX = 0$, since the mean vector does not enter into the following discussion. There exists an orthogonal linear transformation

$$Z = \Phi'X$$

such that the covariance matrix of Z ,

$$E(ZZ') = \Lambda = \begin{bmatrix} \lambda_1 & & 0 \\ & \ddots & \\ 0 & & \lambda_p \end{bmatrix},$$

where $\lambda_1 \geq \dots \geq \lambda_p$ are the roots (eigenvalues) of the characteristic equation $|\Sigma - \lambda I| = 0$, and $\Phi = (\phi_1 \dots \phi_p)$, with ϕ_j the eigenvector corresponding to λ_j . The r th component of Z , $Z_r = \phi_r'X$ is called the r th principal component of X and has maximum variance of all normalized ($\phi_r' \phi_r = 1$) linear combinations uncorrelated with Z_1, \dots, Z_{r-1} . In particular Z_1 the first principal component has maximum variance of any normalized linear

combination. Moreover,

$$\sum_{i=1}^p E(Z_i)^2 = \sum_{i=1}^p \lambda_i = \text{tr}A = \text{tr}E = \sum_{i=1}^p E(x_i)^2 \quad (1.0)$$

Principal components have found application especially in exploratory studies where the number p of variables under consideration may be large. It may happen that a few of the principal components explain most of the variation (as shown by (1.0)) of the original p variables. Another criterion often applied to the use of principal components is that they should have some sort of "reasonable" interpretation in their own right.

When the variables x_1, \dots, x_p are to be used as possible explanatory variables in a linear model, the use of principal components gives the additional advantage that the Z_j 's are mutually uncorrelated. A discussion of the use of principal components in regression analysis is given in [8] along with some examples dealing with economic modelling.

The vectors ϕ_1, \dots, ϕ_p form a basis for the p dimensional Euclidean space R_p . If ψ_1, \dots, ψ_p is any other orthonormal basis for R_p , it can be shown [10 p. 400] that for any $k \leq p$,

$$E\left\{\sum_{i=1}^p \left|x_i - \sum_{v=1}^k Z_v \phi_{iv}\right|^2\right\} \leq E\left\{\sum_{i=1}^p \left|x_i - \sum_{v=1}^k U_v \psi_{iv}\right|^2\right\}, \quad (1.1)$$

$$\text{where } \phi_i = \begin{bmatrix} \phi_{i1} \\ \vdots \\ \phi_{ip} \end{bmatrix}, \quad \psi_i = \begin{bmatrix} \psi_{i1} \\ \vdots \\ \psi_{ip} \end{bmatrix} \quad \text{and } U_v = \psi_v' X, \text{ or}$$

that the k components Z_1, \dots, Z_k give a better approximation in the sense of expected squared error to x_1, \dots, x_p than do the k coefficients of any other orthogonal basis. Moreover, when X is scaled so that $\sum_{i=1}^p E(x_i^2) = 1$, then,

$$-\sum_{i=1}^k \lambda_i \log \lambda_i \leq -\sum_{i=1}^k \rho_i \log \rho_i, \quad (1.2)$$

where $\rho_i = E(U_i^2)$. Equality holds in (1.2) only when

$$\sum_{i=1}^k \lambda_i = \sum_{i=1}^k \rho_i \text{ for every } 0 < k < p. \quad (1.1) \text{ and } (1.2)$$

show that for any fixed $k < p$, the first k principal components carry at least as much of the information contained in the set x_1, \dots, x_p of variables as any k linear combinations, including any subset x_1, \dots, x_k .

When x_1, \dots, x_p are to be considered as possible explanatory variables in a linear model for another variable, say, y , then it is of interest to be able to characterize when the principal components Z_1, \dots, Z_k provide more information relevant to y than some other linear combinations U_1, \dots, U_k .

Let $y, X' = (x_1, \dots, x_p)$ be given random variables

such that $Ey = E(x_i) = 0, i=1, \dots, p; EXX' = \Sigma, EyX' = \Sigma_{12}, \Sigma'_{12} = \Sigma_{21}$. As before let $\lambda_1 \geq \lambda_2 \geq \dots \geq \lambda_p$ and ϕ_1, \dots, ϕ_p be the eigenvalues and corresponding orthonormalized eigenvectors of Σ . Let e_i denote the vector

$$\begin{bmatrix} 0 \\ \vdots \\ 0 \\ 1 \\ 0 \\ \vdots \\ 0 \end{bmatrix} \text{ } i\text{th row}$$

and $\hat{\beta}$ be the regression vector of y on X (i.e. the vector minimizing the expression $E(y - X'\beta)^2$.) Suppose now that y is to be predicted by a linear combination of $k < p$ random variables which are themselves linear combinations of x_1, \dots, x_p . One possibility is to choose some subset x_{i_1}, \dots, x_{i_k} , while another is to choose the first k principal components Z_1, \dots, Z_k .

Let $Z = \begin{bmatrix} z_1 \\ \vdots \\ z_k \end{bmatrix} = \phi^{(k)'} X$, where $\phi^{(k)} = (\phi_1, \dots, \phi_k)$, and

let $W = \begin{bmatrix} w_1 \\ \vdots \\ w_k \end{bmatrix} = E^{(k)'} X$, where $E^{(k)} = (e_{11}, \dots, e_{1k})$.

Let $\hat{y}_Z = Z' \hat{a}$ denote the predictor of y based on Z ,

where $\hat{a} = \begin{bmatrix} \hat{a}_1 \\ \vdots \\ \hat{a}_k \end{bmatrix}$ is the vector minimizing $E(y - Z'a)^2$,

and let $\hat{y}_W = W' \hat{c}$ be the predictor of y based on W , where \hat{c} is the vector minimizing $E(y - W'c)^2$. Then

$$\hat{y}_Z = X' \phi^{(k)} \hat{a} = X' \phi, \text{ while } \hat{y}_W = X' E^{(k)} \hat{c} = X' \psi.$$

Definition: Let u be a vector in R_p and define the norm $||u||$ by $||u||^2 = (u, u)$, where $(u, v) = u' \Gamma v$.

Lemma (1.1): $E(y - \hat{y}_Z)^2 < E(y - \hat{y}_W)^2$ iff $||\hat{\beta} - \psi||$

$$< ||\hat{\beta} - \psi||.$$

Proof: $E(y - \hat{y}_Z)^2 = E(y - X'\phi)^2 = E\{(y - X'\hat{\beta}) + X'(\hat{\beta} - \phi)\}^2 = E(y - X'\hat{\beta})^2 + E[(\hat{\beta} - \phi)'XX'(\hat{\beta} - \phi)]$
 $= E(y - X'\hat{\beta})^2 + ||\hat{\beta} - \phi||^2$. Similarly, $E(y - \hat{y}_W)^2 = E(y - X'\hat{\beta})^2 + ||\hat{\beta} - \psi||^2$ from which the lemma

follows.

Theorem 1.1: $E(y - \hat{y}_Z)^2 < E(y - \hat{y}_W)^2$ iff

$$\Sigma_{12}(\Sigma^{-1} - \sum_{i=1}^k \frac{1}{\lambda_i} \phi_i \phi_i') \Sigma_{21} < \Sigma_{12}(\Sigma^{-1} - \tilde{\Sigma}_k) \Sigma_{21}, \text{ where}$$

$$\tilde{\Sigma}_k = \begin{pmatrix} 0 & \dots & 0 \\ \vdots & & \\ 0 & \dots & \sigma^{11}_{11} & 0 & \dots & \sigma^{11}_{12} & 0 & \dots & \sigma^{11}_{1k} & 0 & \dots & 0 \\ \vdots & & & & & & & & & & & \\ 0 & \dots & \sigma^{12}_{21} & 0 & \dots & \sigma^{12}_{22} & 0 & \dots & \sigma^{12}_{2k} & 0 & \dots & 0 \\ \vdots & & & & & & & & & & & \\ 0 & \dots & \sigma^{1k}_{k1} & 0 & \dots & \sigma^{1k}_{k2} & 0 & \dots & \sigma^{1k}_{kk} & 0 & \dots & 0 \\ \vdots & & & & & & & & & & & \\ 0 & \dots & 0 & \dots & 0 & \dots & 0 & \dots & 0 & \dots & 0 \end{pmatrix},$$

and where σ^{ij}_{jk} is the jk th element of S_k^{-1} where

$$S_k = \begin{pmatrix} \sigma^{11}_{11} & \dots & \sigma^{11}_{1k} \\ \vdots & & \\ \sigma^{1k}_{k1} & \dots & \sigma^{1k}_{kk} \end{pmatrix}.$$

Proof: $||\hat{\beta} - \phi||^2 = (\hat{\beta} - \phi, \hat{\beta} - \phi) = (\hat{\beta}, \hat{\beta}) - 2(\hat{\beta}, \phi)$

$$+ (\phi, \phi). \quad (\hat{\beta}, \hat{\beta}) = \Sigma_{12} \Sigma^{-1} \Sigma_{21}, \quad (\hat{\beta}, \phi) = \hat{\beta}' \Sigma \phi$$

$$= \hat{\beta}' \Sigma \phi^{(k) \hat{a}} = \Sigma_{12} \Sigma^{-1} \Sigma \phi^{(k) \hat{a}} = \Sigma_{12} \phi^{(k) \hat{a}}. \quad \text{But,}$$

$$\hat{a} = [E(ZZ')]^{-1} E(Zy) =$$

$$[E(\phi^{(k)'} X X' \phi^{(k)})]^{-1} E(\phi^{(k)'} X y)$$

$$= [\phi^{(k)'} \Sigma \phi^{(k)}]^{-1} \phi^{(k)'} \Sigma_{21}$$

$$= \left[\begin{bmatrix} \phi_1' \\ \vdots \\ \phi_k' \end{bmatrix} (\Sigma \phi_1 \dots \Sigma \phi_k) \right]^{-1} \phi^{(k)'} \Sigma_{21}$$

$$= \begin{bmatrix} \lambda_1 & & 0 \\ & \ddots & \\ 0 & & \lambda_k \end{bmatrix}^{-1} \phi^{(k)'} \Sigma_{21} = \Lambda^{-1} \phi^{(k)'} \Sigma_{21},$$

so that, $(\hat{\beta}, \phi) = \Sigma_{12} \phi^{(k)'} \Lambda^{-1} \phi^{(k)'} \Sigma_{21}.$

$$(\phi, \phi) = \phi \Sigma \phi = \hat{a}' \phi^{(k)'} \Sigma \phi^{(k) \hat{a}}$$

$$= \Sigma_{12} \phi^{(k)'} \Lambda^{-1} \phi^{(k)'} \Sigma \phi^{(k) \hat{a}} \Lambda^{-1} \phi^{(k)'} \Sigma_{21}$$

$$= \Sigma_{12} \phi^{(k)'} \Lambda^{-1} \Lambda \Lambda^{-1} \phi^{(k)'} \Sigma_{21} = \Sigma_{12} \phi^{(k)'} \Lambda^{-1} \phi^{(k)'} \Sigma_{21}.$$

$$\text{Hence, } ||\hat{\beta} - \phi||^2 = \Sigma_{12} \Sigma^{-1} \Sigma_{21} - \Sigma_{12} \phi^{(k)} \Lambda_{-1}^{-1} \phi^{(k)'} \Sigma_{21}$$

$$= \Sigma_{12} \Sigma^{-1} \Sigma_{21} - \Sigma_{12} \left(\sum_{i=1}^k \frac{1}{\lambda_i} \phi_i \phi_i' \right) \Sigma_{21}$$

$$= \Sigma_{12} \left(\Sigma^{-1} - \sum_{i=1}^k \frac{1}{\lambda_i} \phi_i \phi_i' \right) \Sigma_{21}.$$

$$\text{Similarly, } ||\hat{\beta} - \psi||^2 = (\hat{\beta}, \hat{\beta}) - 2(\hat{\beta}, \psi) + (\psi, \psi).$$

$$(\hat{\beta}, \psi) = \hat{\beta}' \Sigma E^{(k)} \hat{c}, \text{ where } \hat{c} = [E(WW')]^{-1} E(Wy)$$

$$= [E(E^{(k)'} X X' E^{(k)})]^{-1} E(E^{(k)'} X y)$$

$$= [E^{(k)'} \Sigma E^{(k)}]^{-1} E^{(k)'} \Sigma_{21}$$

$$= \begin{bmatrix} e_{11}' \\ \vdots \\ e_{1k}' \end{bmatrix} (\Sigma e_{11} \dots \Sigma e_{1k})^{-1} E^{(k)'} \Sigma_{21}$$

$$= \begin{bmatrix} \sigma_{11}^{11} & \dots & \sigma_{11}^{1k} \\ \vdots & & \vdots \\ \sigma_{1k}^{11} & \dots & \sigma_{1k}^{1k} \end{bmatrix}^{-1} E^{(k)'} \Sigma_{21} = S_k^{-1} E^{(k)'} \Sigma_{21}.$$

$$\text{Thus, } (\hat{\beta}, \psi) = \Sigma_{12} \Sigma^{-1} \Sigma E^{(k)} S_k^{-1} E^{(k)'} \Sigma_{21}.$$

$$= \Sigma_{12} E^{(k)} S_k^{-1} E^{(k)'} \Sigma_{21}, \text{ while } (\psi, \psi) = \psi' \Sigma \psi$$

$$\begin{aligned}
&= \Sigma_{12} E^{(k)} S_k^{-1} E^{(k)'} \Sigma E^{(k)} S_k^{-1} E^{(k)'} \Sigma_{21} \\
&= \Sigma_{12} E^{(k)} S_k^{-1} S_k S_k^{-1} E^{(k)'} \Sigma_{21} = \Sigma_{12} E^{(k)} S_k^{-1} E^{(k)'} \Sigma_{21}.
\end{aligned}$$

$$\text{Hence } ||\hat{\beta} - \psi||^2 = \Sigma_{12} \Sigma^{-1} \Sigma_{21} - \Sigma_{12} E^{(k)} S_k^{-1} E^{(k)'} \Sigma_{21}$$

$$= \Sigma_{12} [\Sigma^{-1} - \tilde{\Sigma}_k] \Sigma_{21},$$

and the theorem follows from lemma 1.1.

In practice if x_1, \dots, x_p are highly inter-correlated, then it might be expected that $\sum_{i=1}^k \frac{1}{\lambda_i} \phi_i \phi_i'$ would more closely approximate Σ^{-1} than would $\tilde{\Sigma}_k$. If at the same time, y is essentially equally correlated with all or many of the x_j 's, then it would seem likely that the inequality would hold. In particular, this situation might arise when x_1, \dots, x_p are elements of a time series $x(t), x(t-1), \dots, x(t-p+1)$ and $y = y(t)$.

As an example, suppose $x(t)$ is a first order autoregressive process with $x(t) - .9x(t-1) = Z(t)$ (a white noise process). Then the covariance matrix Σ of $x(t), x(t-1), x(t-2)$ can be taken as

$$\Sigma = \begin{bmatrix} 1.0 & .9 & .81 \\ .9 & 1.0 & .9 \\ .81 & .9 & 1.0 \end{bmatrix}, \text{ with eigenvalues } \lambda_1 = 2.74,$$

$$\lambda_2 = .19, \lambda_3 = .07 \text{ and eigenvectors}$$

$$\phi_1 = \begin{bmatrix} .571 \\ .590 \\ .571 \end{bmatrix}, \quad \phi_2 = \begin{bmatrix} -.707 \\ 0.0 \\ .707 \end{bmatrix}, \quad \phi_3 = \begin{bmatrix} -.417 \\ .807 \\ -.417 \end{bmatrix}.$$

$$\Sigma^{-1} = \begin{bmatrix} 5.232 & -4.687 & -.027 \\ -4.687 & 9.431 & -4.687 \\ -.027 & -4.687 & 5.232 \end{bmatrix} \quad . \quad \text{Letting } k = 1,$$

$$\frac{1}{\lambda_1} \phi_1 \phi_1' = \begin{bmatrix} .117 & .121 & .117 \\ .121 & .125 & .121 \\ .117 & .121 & .117 \end{bmatrix}, \quad \text{and}$$

$$\tilde{\Sigma}_1 = \begin{bmatrix} 1 & 0 & 0 \\ 0 & 0 & 0 \\ 0 & 0 & 0 \end{bmatrix} \quad . \quad \text{Letting } \Sigma_{12} = (1 \quad .99 \quad .98),$$

$$\Sigma_{12}(\Sigma^{-1} - \frac{1}{\lambda_1} \phi_1 \phi_1')\Sigma_{21} = .017 \text{ while}$$

$$\Sigma_{12}(\Sigma^{-1} - \tilde{\Sigma}_1)\Sigma_{21} = .071.$$

CHAPTER II

STOCHASTIC PROCESSES AND REPRESENTATIONS

The type of stochastic process $x(t)$ to be considered will be a family of random variables with index set T , where unless otherwise stated T will denote a countable set of the real line, so that $x(t)$ is a discrete process. The random variables $x(t)$ may be real or complex valued, but will always be assumed to satisfy the condition $E|x(t)|^2 < \infty$ for all $t \in T$. The mean value function of $x(t)$ will be denoted by $\mu(t)$, and unless otherwise stated it will be assumed that $\mu(t) \equiv 0$. The covariance function will be denoted by $r(s, t) = E\{x(s) \overline{x(t)}\}$ where the bar denotes complex conjugate. Of particular importance is the case where $x(t)$ is a second order stationary process. In that case the covariance function has the form

$$E\{x(s) \overline{x(s-t)}\} = r(t),$$

and $r(t)$ has the spectral representation

$$r(t) = \int_{-\pi}^{\pi} e^{it\omega} dF(\omega),$$

where $F(\omega)$, the spectral distribution function of $x(t)$, is real valued, nondecreasing, bounded, and can be chosen so that $F(-\infty) = 0$, $F(+\infty) = r(0)$, and $F(\omega)$ is continuous to the right. If $F(\omega)$ is absolutely continuous then $f(\omega) = F'(\omega)$ is the spectral density function of $x(t)$. In addition, if $x(t)$ is second order stationary, it has the spectral representation

$$x(t) = \int_{-\pi}^{\pi} e^{it\omega} d\xi(\omega),$$

where $\xi(\omega)$ is a process of orthogonal increments which can be chosen so that $E \xi(\omega) = 0$, $E|\xi(\omega)|^2 = F(\omega)$, and $E|d\xi(\omega)|^2 = dF(\omega)$. ($\xi(\omega)$ is called the spectral process associated with $x(t)$.) From this spectral representation the process $x(t)$ can be thought of as being built up of mutually orthogonal elementary harmonic oscillations $e^{it\omega} d\xi(\omega)$.

Let $x(t)$ be a second order stationary process and consider a process $y(t)$ formed from $x(t)$ by a linear transformation of the type

$$y(t) = \sum_{v=1}^n b_v x(t - t_v). \quad (2.1)$$

Then

$$\begin{aligned} y(t) &= \sum_{v=1}^n b_v \int_{-\infty}^{\infty} e^{i(t-t_v)\omega} d\xi_x(\omega) \\ &= \int_{-\infty}^{\infty} e^{it\omega} \left(\sum_{v=1}^n b_v e^{-it_v\omega} \right) d\xi_x(\omega). \end{aligned}$$

Thus, $y(t)$ is a second order stationary process with elementary harmonic oscillations $h(\omega) d\xi_x(\omega)$, where $h(\omega)$

$$= \sum_{v=1}^n b_v e^{-it_v\omega}. \quad \text{The covariance function } r_y(t) \text{ of } y(t)$$

has the representation

$$r_y(t) = \int_{-\infty}^{\infty} e^{it\omega} |h(\omega)|^2 dF_x(\omega).$$

If $x(t)$ has spectral density function $f_x(\omega)$ then $y(t)$ has spectral density function $f_y(\omega) = |h(\omega)|^2 f_x(\omega)$.

Thus for a given frequency ω , $f_x(\omega)$ is multiplied by

the amount $|h(\omega)|^2$ to give $f_y(\omega)$. This may have the effect of greatly repressing certain frequencies while enhancing others. For this reason the linear operation (2.1) may be thought of as a filter applied to the process

$x(t)$. $|h(\omega)|^2$ is called the transfer function of the filter and $h(\omega)$ the gain of the filter.

Karhunen - Loève Representations

Theorem [see 7 p. 478]

Let $x(t)$ be a stochastic process continuous in quadratic mean and defined for $t \in T$ a closed interval. Then $x(t)$ has an orthogonal representation of the form

$$x(t) = \sum_{n=1}^{\infty} \psi_n(t) Z_n,$$

where $\psi_n(t)$ is an eigenfunction of the covariance function $R(t, t') = E\{x(t) x(t')\}$. That is

$$\int_T R(t, t') \psi_n(t') dt' = \lambda_n \psi_n(t),$$

and

$\int_T \psi_m(t) \overline{\psi_n(t)} dt = \delta_{mn}$. The random variables Z_n are given by

$$Z_n = \int_T x(t) \overline{\psi_n(t)} dt,$$

and the Z_n 's have the property that $E(Z_m \overline{Z_n}) = \lambda_n \delta_{mn}$.

When $x(t)$ is a finite discrete process, T can be taken as the set $T = \{1, 2, \dots, N\}$, and $R(t, t')$ becomes the

$N \times N$ matrix

$$R^{(N)} = \begin{bmatrix} R(1,1) & \dots & R(1,N) \\ R(2,1) & \dots & R(2,N) \\ \vdots & & \vdots \\ R(N,1) & \dots & R(N,N) \end{bmatrix},$$

with eigenvalues $\lambda_1 \geq \lambda_2 \geq \dots \geq \lambda_N$, and corresponding orthonormal set of eigenvectors ψ_1, \dots, ψ_N , where

$$\psi_k = \begin{bmatrix} \psi_k(1) \\ \vdots \\ \psi_k(N) \end{bmatrix}. \quad \text{The representation for } x(t) \text{ then}$$

becomes

$$x(t) = \sum_{n=1}^N \psi_n(t) Z_n,$$

where

$$Z_n = \sum_{t=1}^N x(t) \overline{\psi_n(t)}.$$

Thus in this case Z_n is the n th principal component of the random variables $x(1), \dots, x(N)$.

Principal Component Processes

Let $x(t)$, $t = 1, 2, \dots$ be a discrete stochastic process.

Definition: $x(t)$ will be called covariance stationary if $E\{(x(t) - \mu(t))(x(s) - \mu(s))\} = r(t - s)$, where

$\mu(t) = E x(t)$. ($\mu(t)$ is not necessarily constant.) For any fixed positive integer n let $R^{(n)}$ be the covariance matrix

$$R^{(n)} = \begin{bmatrix} r(0) & r(1) & \dots & r(n) \\ r(-1) & r(0) & \dots & r(n-1) \\ \vdots & \vdots & \ddots & \vdots \\ r(-n) & r(-n+1) & \dots & r(0) \end{bmatrix},$$

$\lambda_0^{(n)} \geq \lambda_1^{(n)} \geq \lambda_2^{(n)} \geq \dots \geq \lambda_n^{(n)}$ be the eigenvalues of $R^{(n)}$, and $\phi_0^{(n)}, \phi_1^{(n)}, \dots, \phi_n^{(n)}$ be the corresponding set of orthonormalized eigenvectors.

Definition: The processes $Z_j^{(n)}(t); j = 0, 1, \dots, n$ where

$$Z_j^{(n)}(t) = \sum_{v=0}^n \phi_j^{(n)}(v) x(t-v)$$

will be called the principal component processes of $x(t)$.

From this definition and the usual properties of principal components it follows that:

(1) $Z_0(t)$ is that linear combination

$$\sum_{i=0}^n a_{i0} x(t-i) \text{ with maximum variance sub-}$$

ject to the restriction that $\sum_{i=1}^n a_{i0}^2 = 1$.

$Z_1^{(n)}(t)$ is the linear combination

$$\sum_{i=1}^n a_{i1} x(t-i) \text{ with maximum variance}$$

subject to the restrictions that $\sum_{i=1}^n a_{i1}^2 = 1$

and $\sum_{i=1}^n a_{i0} \overline{a_{i1}} = 0$, and so forth for

$z_j^{(n)}(t)$, $j = 0, \dots, n$.

$$(ii) \quad E\{z_1^{(n)}(t) \overline{z_j^{(n)}(t)}\} = \lambda_1 \delta_{1j}.$$

(iii) $x(t)$ has the representation,

$$x(t) = \sum_{v=0}^n \overline{\phi_v(0)} z_v^{(n)}(t).$$

This representation differs in general from the Karhunen-Loève representation or the spectral representation in that the index t is contained in the random element. Thus the representation is in terms of constants and random processes rather than random variables and deterministic functions of t .

$$(iv) \quad r(0) = \sum_{v=0}^n |\phi_v(0)|^2 \lambda_v.$$

An interpretation of these processes can be seen more readily in the case that $x(t)$ is a second order

stationary process. Let $x(t)$ be a discrete stationary process with spectral process $\xi_x(\omega)$, and assume moreover, that $x(t)$ has a spectral density function $f_x(\omega)$. Then

$$z_j^{(n)}(t) = \sum_{v=0}^n \phi_j^{(n)}(v) x(t-v), \quad j = 0, \dots, n$$

is a linear operation on $x(t)$ and hence, it follows that,

(v) $z_j^{(n)}(t)$ is a stationary process,

(vi) $z_j^{(n)}(t)$ is built up of elementary harmonic oscillations of the form $e^{it\omega} g_v^{(n)}(\omega) d\xi_x(\omega)$ where $g_v^{(n)}(\omega) = \sum_{k=0}^n e^{-ik\omega} \phi_v^{(n)}(k)$,

(vii) $z_j^{(n)}(t)$ has spectral density function

$$f_{z_j^{(n)}}(\omega) = |g_j^{(n)}(\omega)|^2 f_x(\omega).$$

From (iii), (vi), (vii) it follows that the spectral representations for $x(t)$ and $r(0)$ may be written in the

forms

$$\begin{aligned}
 x(t) &= \int_{-\pi}^{\pi} e^{it\omega} d\xi_x(\omega) \\
 &= \int_{-\pi}^{\pi} e^{it\omega} \overline{\phi_0(0)} g_0(\omega) d\xi_x(\omega) + \dots \\
 &\quad + \overline{\phi_n(0)} g_n(\omega) d\xi_x(\omega),
 \end{aligned}$$

and

$$\begin{aligned}
 r(0) &= \int_{-\pi}^{\pi} f_x(\omega) d\omega \\
 &= \int_{-\pi}^{\pi} (|\phi_0(0)|^2 |g_0(\omega)|^2 f_x(\omega) + \dots \\
 &\quad + |\phi_n(0)|^2 |g_n(\omega)|^2 f_x(\omega)) d\omega
 \end{aligned}$$

which gives a decomposition for the elementary harmonic oscillations $e^{it\omega} d\xi_x(\omega)$ and the spectral density function $f_x(\omega)$. From the properties (i) and (ii) it follows that

$$\int_{-\pi}^{\pi} |g_0(\omega)|^2 f_x(\omega) d\omega = \omega_0 \geq \int_{-\pi}^{\pi} |h(\omega)|^2 f_x(\omega) d\omega,$$

for any $h(\omega)$ satisfying

$$h(\omega) = \sum_{v=0}^n b_v e^{-iv\omega}, \quad \sum_{v=0}^n b_v^2 = 1.$$

Thus, the filter function $g_0(\omega)$ passes the most important portion of the spectrum of $x(t)$, while $g_1(\omega), \dots, g_n(\omega)$ pass increasingly smaller portions of $f_x(\omega)$. The exact properties of these filter functions are determined by the filter coefficients

$$\phi_v(k), \quad k = 0, \dots, n; \quad v = 0, \dots, n$$

which in turn are determined by the covariance structure of the process $x(t)$. These properties are further illustrated by the following special case.

$x(t)$ A Periodic Process

Let $x(t)$ be a stationary process such that $x(t + N) = x(t)$ for some integer N . This process can be thought of as containing only a finite number of elementary harmonic oscillations and, hence, has the representation

$$x(t) = \sum_{j=0}^{N-1} e^{i\left(\frac{2\pi j}{N}\right)t} x_j, \quad (2.2)$$

where the

$$x_j = \frac{1}{N} \sum_{k=0}^{N-1} e^{-i\left(\frac{2\pi j}{N}\right)k} x(k), \quad j = 0, \dots, n-1 \text{ are}$$

mutually orthogonal random variables. (2.2) can be put in the more usual form by letting

$$\omega_j = \begin{cases} \frac{2\pi j}{N}, & j \leq \frac{N}{2} \\ \frac{2\pi j}{N} - 2\pi, & j > \frac{N}{2} \end{cases}.$$

Then (2.2) becomes

$$x(t) = \sum_{j=0}^{N-1} e^{i\omega_j t} x_j, \quad -\pi \leq \omega_j \leq \pi.$$

From the definition of $x(t)$, $E x(t) \overline{x(t - (N - j))}$
 $= E x(t) \overline{x(t + j)}$ or $r(n - j) = r(-j)$. Hence, the covariance matrix

$$R^{(N-1)} = \begin{bmatrix} r(0) & r(1) & \dots & r(N-1) \\ r(-1) & r(0) & \dots & r(N-2) \\ \vdots & \vdots & \ddots & \vdots \\ r(-N+1) & r(-N+2) & \dots & r(0) \end{bmatrix}$$

$$= \begin{bmatrix} r(0) & r(1) & \dots & r(N-1) \\ r(N-1) & r(0) & \dots & r(N-2) \\ \vdots & \vdots & \ddots & \vdots \\ r(1) & r(2) & \dots & r(0) \end{bmatrix}$$

is a circular matrix. As such it has eigenvalues given by

$$\lambda_j = \sum_{k=0}^{N-1} \rho_j^k r(k), \text{ where } \rho_j \text{ is an } N\text{th root of}$$

unity. The corresponding eigenvector is given by

$$\phi_j^{(n)} = \begin{bmatrix} 1 \\ \rho_j \\ \rho_j^2 \\ \vdots \\ \rho_j^{N-1} \end{bmatrix} \frac{1}{\sqrt{N}}$$

Putting $\rho_j = e^{\frac{12\pi j}{N}}$ gives

$$\phi_j^{(n)} = \begin{bmatrix} 1 \\ e^{\frac{12\pi j}{N}} \\ e^{\frac{12\pi 2j}{N}} \\ \vdots \\ e^{\frac{12\pi (N-1)j}{N}} \end{bmatrix} \frac{1}{\sqrt{N}} \quad (2.3)$$

The principal component process $Z_j^{(n)}(t)$ then has the form

$$\begin{aligned} Z_j^{(n)}(t) &= \sum e^{\frac{12\pi j v}{N}} x(t-v) \frac{1}{\sqrt{N}} \\ &= \frac{1}{\sqrt{N}} \sum_{v=0}^{N-1} e^{\frac{12\pi j v}{N}} \sum_{k=0}^{N-1} e^{\frac{12\pi k(t-v)}{N}} x_k \\ &= \frac{1}{\sqrt{N}} \sum_{k=0}^{N-1} \sum_{v=0}^{N-1} e^{\frac{12\pi (1-k)v}{N}} e^{\frac{12\pi k t}{N}} x_k \end{aligned}$$

$$= \frac{1}{\sqrt{N}} \sum_{k=0}^{N-1} N \delta_{jk} e^{i \frac{2\pi k}{N} t} x_k = \sqrt{N} e^{i \frac{2\pi j}{N} t} x_j,$$

so that $Z_j(t)$ consists only of the elementary harmonic oscillation at frequency $\frac{2\pi j}{N}$. This may also be seen from the filter function corresponding to $Z_j(t)$, which is

$$g_j(\omega) = \frac{1}{\sqrt{N}} \sum_{v=0}^{N-1} e^{i \frac{2\pi j}{N} v} e^{-i v \omega}$$

$$= \frac{1}{\sqrt{N}} \sum_{v=0}^{N-1} e^{i \left(\frac{2\pi j}{N} - \omega \right) v}$$

But $x(t)$ only contains oscillations at frequencies

$$\omega_k = \frac{2\pi k}{N}, \quad k=0, \dots, N-1 \text{ and } g_j(\omega_k) = \sqrt{N} \delta_{jk}.$$

The representation (iii) in this case yields

$$x(t) = \sum_{j=0}^{N-1} \frac{1}{\sqrt{N}} \sqrt{N} e^{i \frac{2\pi j}{N} t} x_j$$

$$= \sum_{j=0}^{N-1} e^{i \frac{2\pi j}{N} t} x_j$$

which is identical with the spectral representation (2.2).

On the other hand, since $x(t + N) = x(t)$, the process

$x(t)$ consists of only a finite number of distinct random

variables. Thus, considering only the distinct $x(t)$'s, the

representation (2.2) is identical with the Karhunen representation with $Z_j = x_j$.

Since an arbitrary process may be approximated by a periodic one by allowing the period to tend to infinity, it seems reasonable to inquire whether some of the properties exhibited by the principal component processes for the periodic case might not hold in some limiting sense for arbitrary processes. For instance, the form of the eigenvalues of the circular matrix

$$\lambda_j^{(N-1)} = \sum_{k=0}^{N-1} \left(e^{i \frac{2\pi j}{N} k} \right) r(k)$$

might suggest that the set $\{\lambda_v^{(n)}\}_{v=0}^n$ of eigenvalues of the covariance matrix of a stationary process might in some way approximate its spectral density function.

These ideas are made more precise using the theory of Toeplitz forms developed by Grenander and Szegő [5].

Definition: Let $f(x)$ be a real valued function defined on

$[-\pi, \pi]$ such that $\int_{-\pi}^{\pi} |f(x)| dx$ exists, and let $c^{(n)}$ be

the matrix $(c_{\mu, \nu})$; $\mu, \nu = 0, \dots, n$, where

$$c_k = \frac{1}{2\pi} \int_{-\pi}^{\pi} e^{-ikx} f(x) dx.$$

Then $c^{(n)}$ will be called the Toeplitz matrix of order n associated with the function $f(x)$. Similarly, if $F(x)$ is monotone nondecreasing in $[-\pi, \pi]$, the Toeplitz matrix associated with $F(x)$ is $c^{(n)} = (c_{\nu-\mu})$ where

$$c_k = \frac{1}{2\pi} \int_{-\pi}^{\pi} e^{-ikx} dF(x).$$

The quadratic form $X c^{(n)} X^*$ is called the Toeplitz form associated with $f(x)$ or $F(x)$ where $X = (x_0, x_1, \dots, x_n)$ is an arbitrary complex vector.

Proposition: Let $x(t)$ be a second order stationary stochastic process and $R^{(n)}$ the covariance matrix

$$R^{(n)} = \begin{bmatrix} r(0) & . & . & . & r(n) \\ \vdots & & & & \\ \vdots & & & & \\ r(-n) & . & . & . & r(0) \end{bmatrix}$$

Then $R^{(n)}$ is the Toeplitz matrix associated with the spectral distribution $2\pi F_x(\omega)$ (or spectral density function $2\pi f(\omega)$ if it exists).

Proof: Bochner's theorem for the discrete case gives

$$r(n) = \int_{-\pi}^{\pi} e^{in\omega} dF_x(\omega) \text{ so that}$$

$$r(-n) = \frac{1}{2\pi} \int_{-\pi}^{\pi} e^{-in\omega} d[2\pi F_x(\omega)].$$

Definition: For each n consider the two sets of real

numbers $\{a_v^{(n)}\}_{v=0}^n$ and $\{b_v^{(n)}\}_{v=0}^1$, where there exists a

constant K independent of n and v , such that $|a_v^{(n)}| < K$,

$|b_v^{(n)}| < K$. $\{a_v^{(n)}\}$ and $\{b_v^{(n)}\}$, $n \rightarrow \infty$ are said to be equally distributed in $[-K, K]$ if

$$\lim_{n \rightarrow \infty} \frac{1}{n+1} \sum_{v=0}^n [F(a_v^{(n)}) - F(b_v^{(n)})] = 0,$$

where $F(t)$ is an arbitrary continuous function on $[-K, K]$.

Theorem: Let $f(x)$ be a real valued function such that

$\int_{-\pi}^{\pi} |f(x)| dx < \infty$, and let m and M denote the essential

lower and upper bounds (assumed finite) respectively of $f(x)$. If $F(\lambda)$ is any continuous function on $[m, M]$, then

$$\lim_{n \rightarrow \infty} \frac{F(\lambda_0^{(n)}) + \dots + F(\lambda_n^{(n)})}{n+1} = \frac{1}{2\pi} \int_{-\pi}^{\pi} F(f(x)) dx,$$

(2.4)

where $\lambda_0^{(n)} \geq \dots \geq \lambda_n^{(n)}$ are the eigenvalues of the

matrix $c^{(n)}$ associated with $f(x)$. For proof see [5]. By considering the integral on the right hand side of (2.4) as a limit of approximating sums, the theorem shows that the sets $\{\lambda_v^{(n)}\}_{v=0}^n$ and $\{f(\pi - \frac{2(v+1)\pi}{n+2})\}_{v=0}^n$ are equally distributed as $n \rightarrow \infty$, and by proper choice of $P(\lambda)$ it follows that

$$\lim_{n \rightarrow \infty} \lambda_0^{(n)} = M. \quad (2.5)$$

These results show that when $x(t)$ is a second order stationary process with spectral density function $f_x(\omega)$, the eigenvalues $\lambda_0^{(n)}, \dots, \lambda_n^{(n)}$ of $R^{(n)}$ tend to approximate $2\pi f_x(\omega)$ at an equally spaced set of points in $[-\pi, \pi]$, and in particular,

$$\lim_{n \rightarrow \infty} \lambda_0^{(n)} = \max_{-\pi \leq \omega \leq \pi} 2\pi f_x(\omega)$$

Pertaining to the asymptotic behavior of eigenvectors, it is shown in [5] that a Toeplitz matrix can be approximated according to a certain norm by a hermitian matrix (with known eigenvalues and eigenvectors) in such a way that the distributions of eigenvalues of the two matrices are asymptotically equal. The mode of convergence, however, does not seem to be strong enough to guarantee the convergence of a transfer function given by an eigenvector

of a Toeplitz matrix to the same limit as that of the transfer function given by the corresponding eigenvector of the approximating matrix. In particular, it is shown that an arbitrary Toeplitz matrix can be approximated by a matrix with eigenvectors given by (2.3). This might suggest that the principal component processes might, in general, tend to give a frequency decomposition as in the case of the periodic process. That this is not always the case can be seen by considering a process $x(t)$ of the opposite extreme.

$x(t)$ A Discrete White Noise Process

Let $x(t)$ be such that $E x(t) = 0$, and $E\{x(t) \overline{x(t')}\} = \sigma^2 \delta_{tt'}$. The spectral density function of this process is

$$f_x(\lambda) = \frac{\sigma^2}{2\pi}, \quad -\pi \leq \lambda \leq \pi,$$

and the covariance matrix

$$R(n) = \begin{bmatrix} \sigma^2 & & & 0 \\ & \ddots & & \\ & & \ddots & \\ 0 & & & \sigma^2 \end{bmatrix},$$

with eigenvalues $\lambda_0^{(n)} = \dots = \lambda_n^{(n)} = \sigma^2$ and corresponding eigenvectors $\phi_0^{(n)}, \dots, \phi_n^{(n)}$ which may be taken as

$$\begin{pmatrix} 1 \\ 0 \\ \vdots \\ 0 \end{pmatrix}, \begin{pmatrix} 0 \\ 1 \\ \vdots \\ 0 \end{pmatrix}, \dots, \begin{pmatrix} 0 \\ \vdots \\ 0 \\ 1 \end{pmatrix}. \quad \text{The transfer function}$$

given by $\phi_k^{(n)}$ is hence,

$$|\phi_k^{(n)}(\lambda)|^2 = |e^{-ik\lambda}|^2 = 1,$$

so that each transfer function passes each frequency with equal weighting. Thus, while the eigenvectors in this case are not uniquely determined (and hence neither are the principal component processes) for at least one choice of eigenvectors the principal component processes do not provide a frequency decomposition of the original process.

It will be seen in what follows that for many other common processes certain of the associated principal component processes tend to give meaningful frequency decompositions.

x(t) A First Order Moving Average Process

Let $Z(t)$ be a white noise process satisfying $E Z(t) = 0$, $E\{Z(t), Z(t')\} = \delta_{tt'}$, and let $x(t) = Z(t) + \rho Z(t-1)$, where $0 < |\rho| < 1$.

Then

$$R^{(n)} = \begin{bmatrix} 1+\rho^2 & \rho & 0 & \vdots & \vdots & \vdots & \vdots & 0 \\ \rho & 1+\rho^2 & \rho & 0 & \vdots & \vdots & \vdots & 0 \\ 0 & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots \\ \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots \\ \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots \\ 0 & \dots & 0 & \rho & 1+\rho^2 & \vdots & \vdots & \vdots \end{bmatrix}.$$

The spectral density function $f_x(\omega)$ of $x(t)$ is

$$\begin{aligned} f_x(\omega) &= \frac{1}{2\pi} |1 + \rho^2 + \rho e^{-i\omega}| \\ &= \frac{1}{2\pi} (1 + \rho^2 + 2\rho \cos \omega). \end{aligned} \quad (2.6)$$

From the equation of "mechanical quadrature" (see [5])

$$\frac{1}{2\pi} \int_{-\pi}^{\pi} \phi(x) dx = \frac{1}{2N} \sum_{v=-N+1}^N \phi\left(\frac{v\pi}{N}\right),$$

which holds when $\phi(x)$ is a trigonometric polynomial whose degree does not exceed $2N-1$, it can be shown that the eigenvalues of $R^{(n)}$ are given by

$$\lambda_v^{(n)} = (1 + \rho^2) + 2\rho \cos \frac{(v+1)\pi}{n+2} \quad (2.7)$$

with corresponding eigenvectors,

$$\Phi_v^{(n)} = \begin{bmatrix} \phi_{00}^{(n)} \\ \vdots \\ \phi_{vn}^{(n)} \end{bmatrix}, \text{ where}$$

$$\phi_{vk}^{(n)} = \sqrt{\frac{2}{n+2}} \sin \frac{(k+1)(v+1)}{n+2}, \quad v=0, \dots, n. \quad (2.8)$$

When $\rho > 0$, $\lambda_0^{(n)} > \lambda_1^{(n)} > \dots > \lambda_n^{(n)}$ and $\lambda_0^{(n)} \xrightarrow{n \rightarrow \infty} (1+\rho)^2$

$= f_x(0) = \max_{\omega} f_x(\omega)$. If $\rho < 0$, the subscripts on the $\lambda^{(n)}$'s must be reversed in order to have the proper ordering. After relabeling

$$\begin{aligned} \lambda_0^{(n)} &= (1 + \rho^2) + 2\rho \cos \frac{(n+1)\pi}{n+2} \xrightarrow{n \rightarrow \infty} (1 - \rho)^2 \\ &= f_x(\pm\pi) = \max_{\omega} f_x(\omega). \end{aligned}$$

For the case $n = 1$, $\rho > 0$,

$$\lambda_0^{(1)} = (1+\rho^2) + 2\rho \cos \pi/3 > \lambda_1^{(1)} = (1+\rho^2) + 2\rho \cos 2\pi/3,$$

$$\Phi_0^{(1)} = \begin{bmatrix} 1/\sqrt{2} \\ 1/\sqrt{2} \end{bmatrix}, \quad \Phi_1^{(1)} = \begin{bmatrix} 1/\sqrt{2} \\ -1/\sqrt{2} \end{bmatrix},$$

$$Z_0^{(1)}(t) = \frac{1}{\sqrt{2}} x(t) + \frac{1}{\sqrt{2}} x(t-1),$$

$$z_1^{(1)}(t) = \frac{1}{\sqrt{2}} x(t) - \frac{1}{\sqrt{2}} x(t-1).$$

The transfer functions corresponding to $z_0^{(1)}(t)$ and $z_1^{(1)}(t)$ are

$$|g_{z_0^{(1)}}(\omega)|^2 = \left| \frac{1}{\sqrt{2}} + \frac{1}{\sqrt{2}} e^{-i\omega} \right|^2 = 1 + \cos \omega \text{ and}$$

$$|g_{z_1^{(1)}}(\omega)|^2 = \left| \frac{1}{\sqrt{2}} - \frac{1}{\sqrt{2}} e^{-i\omega} \right|^2 = 1 - \cos \omega \text{ as shown}$$

in figures 1 and 2



Figure 1

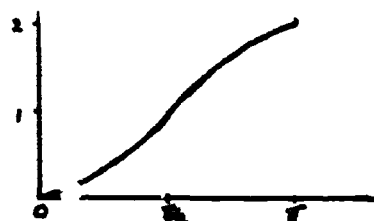


Figure 2

Similarly for any finite n , the eigenvector corresponding to the largest eigenvalue yields a principal component process whose transfer function passes th largest portion of $f_x(\omega)$.

Proposition: $\lim_{n \rightarrow \infty} f_{z_0^{(n)}}(\omega) = 0$ for any fixed $\omega \neq 0$, $\rho > 0$

Proof: $g_{z_0(n)}(\omega) = \sum_{k=0}^n \sqrt{\frac{2}{n+2}} \left(\sin \frac{(k+1)\pi}{n+2} \right) e^{-ik\omega}$

$$= \frac{1}{2i} \sqrt{\frac{2}{n+2}} \sum_{k=0}^n \left[e^{i \frac{(k+1)\pi}{n+2}} - e^{-i \frac{(k+1)\pi}{n+2}} \right] e^{-ik\omega}$$

$$= \frac{1}{2i} \sqrt{\frac{2}{n+2}} \left\{ e^{i \frac{\pi}{n+2}} \sum_{k=0}^n e^{i \left(\frac{\pi}{n+2} - \omega \right) k} - e^{-i \frac{\pi}{n+2}} \sum_{k=0}^n e^{-i \left(\frac{\pi}{n+2} + \omega \right) k} \right\}$$

$$= \frac{1}{2i} \sqrt{\frac{2}{n+2}} \left\{ e^{i \frac{\pi}{n+2}} \frac{e^{i \left(\frac{\pi}{n+2} - \omega \right) (n+1)} - 1}{e^{i \frac{\pi}{n+2} - \omega} - 1} - e^{-i \frac{\pi}{n+2}} \frac{e^{-i \left(\frac{\pi}{n+2} + \omega \right) (n+1)} - 1}{e^{-i \frac{\pi}{n+2} + \omega} - 1} \right\}$$

$$= \frac{1}{2i} \sqrt{\frac{2}{n+2}} \frac{N_{1n}(\omega)}{D_{1n}(\omega)} - \frac{N_{2n}(\omega)}{D_{2n}(\omega)} \quad \text{For } \omega = \omega_0 \neq 0,$$

$D_{1n}(\omega)$ and $D_{2n}(\omega) \rightarrow e^{-i\omega} - 1 \neq 0$, while $|N_{2n}(\omega)| \leq 2$ and $|N_{1n}(\omega)| \leq 2$. Thus as $n \rightarrow \infty$, $g_{z_0(n)}(\omega) \rightarrow 0$ and $f_{z_0(n)}(\omega)$

$|g_{z_0(n)}(\omega)|^2 f_x(\omega) \rightarrow 0$. Hence, the transfer function

$|g_{z_0(n)}(\omega)|^2$ acts like a δ function, $\delta(\omega)$, as $n \rightarrow \infty$. A

similar argument shows that when $\rho < 0$,

$$|g_{z_0(n)}(\omega)|^2 \rightarrow \delta(|\omega| - \pi) \text{ as } n \rightarrow \infty.$$

x(t) A First Order Autoregressive Process

Let $Z(t)$ be a white noise process as above and

$$x(t) - \rho x(t-1) = Z(t). \quad (2.8)$$

(2.8) may also be written

$$(1 - \rho B)x(t) = Z(t),$$

where B represents the "backwards lag operator"

(i.e. $Bx(t) = x(t-1)$), or

$$x(t) = \frac{1}{1 - \rho B} Z(t). \quad (2.9)$$

If $x(t)$ is to be a stationary process then $|\rho| < 1$, in which case (2.9) may formally be written as

$$x(t) = \sum_{v=0}^{\infty} \rho^v B^v Z(t), \text{ or}$$

$$x(t) = \sum_{v=0}^{\infty} \rho^v Z(t-v). \quad (2.10)$$

Using (2.10), $R^{(n)}$ can be seen to have the form

$$R^{(n)} = \sigma^2 \begin{pmatrix} 1 & \rho & \rho^2 & \dots & \rho^{n-1} \\ \rho & 1 & & & \rho^{n-1} \\ \vdots & & & & \\ \vdots & & & & \\ \rho^n & \rho^{n-1} & \dots & \dots & \rho^1 \end{pmatrix}, \quad (2.11)$$

where $\sigma^2 = \sum_{v=0}^{\infty} \rho^{2v}$. $x(t)$ has spectral density function

$$f(\omega) = \frac{1}{2\pi} \left| \frac{1}{1 - e^{i\omega}} \right|^2 = \frac{1}{2\pi} (1 + \rho^2 - 2\rho \cos \omega)^{-1} \quad (2.12)$$

which assumes its maximum at $\omega = 0$ when $\rho > 0$. While explicit expressions for the eigenvalues and eigenvectors of $R^{(n)}$ are unfeasible, something can be said about the asymptotic properties of the largest eigenvalue and corresponding eigenvector.

Definition: A matrix $A = (a_{ij})$ is called positive if $a_{ij} > 0$ for all i, j .

Then when $\rho > 0$, $R^{(n)}$ is a positive matrix. This will be assumed to be the case in what follows.

Theorem (Perron): [see 2 p. 278] If A is a positive matrix, there is a unique eigenvalue $\lambda(A)$ which has greatest absolute value, is positive, and simple. Its associated eigenvector can be taken to be positive (have all positive elements).

Theorem (2.0): [2] $v = \lim_{k \rightarrow \infty} \frac{A^k c}{[\lambda(A)]^k}$ exists and is an eigenvector of A associated with $\lambda(A)$ for any positive vector c . v is unique up to a scalar multiple depending on the choice of c , but otherwise, independent of the choice of c .

Lemma (2.1): Let $\phi_0 = \begin{pmatrix} \phi_{00} \\ \vdots \\ \phi_{0n} \end{pmatrix}$ be the positive eigen-

vector of $R^{(n)}$ corresponding to the largest eigenvalue.

Then $0 < \phi_{00} \leq \phi_{01} \leq \dots \leq \phi_{0 \frac{n-1}{2}} = \phi_{0 \frac{n+1}{2}} \geq \dots \geq \phi_{0n}$

$= \phi_{00}$ if n is odd.

$0 < \phi_{00} \leq \phi_{01} \leq \dots \leq \phi_{0 \frac{n}{2}-1} \leq \phi_{\frac{n}{2}} \geq \phi_{\frac{n}{2}+1} = \phi_{\frac{n}{2}-1}$

$\geq \dots \geq \phi_{0n} = \phi_{00}$ if n is even.

Proof: Only the case n even will be treated explicitly. Positiveness can be assumed as a result of Perron's theorem. From theorem (2.0), if c is chosen to be the vector of 1's, then for any r

$$(R^{(n)})^r c = \begin{bmatrix} S_0^{(r)} \\ S_1^{(r)} \\ \vdots \\ S_n^{(r)} \end{bmatrix},$$

where $S_j^{(r)}$ is the j th row sum of $(R^{(n)})^r$ or $S_j^{(r)} = \sum_{v=0}^n b_{jv}^{(n)} S_v^{(r-1)}$, where $b_{jv}^{(n)}$ is the element in the j th row, v th column of $R^{(n)}$. The equality of $S_j^{(r)}$ and $S_{n-j}^{(r)}$ is obvious from the form of $R^{(n)}$ and c . For

$r = 1$,

$$S_0^{(1)} = 1 + \rho + \rho^2 + \dots + \rho^n < S_1^{(1)} = 1 + 2\rho + \rho^2$$

$$+ \dots + \rho^{n-1} < \dots < S_{\frac{n}{2}}^{(1)} = 1 + 2\rho + 2\rho^2 + \dots + 2\rho^{\frac{n}{2}}$$

$$> \dots > S_n^{(1)} = S_0^{(1)}. \text{ For } r = 2,$$

$$S_0^{(2)} = S_0^{(1)} + \rho S_1^{(1)} + \rho^2 S_2^{(1)} + \dots + \rho^n S_n^{(1)} \text{ while}$$

$$S_1^{(2)} = \rho S_0^{(1)} + S_1^{(1)} + \rho S_2^{(1)} + \dots + \rho^{n-1} S_n^{(1)}, \text{ so}$$

$$S_1^{(2)} - S_0^{(2)} = (1-\rho)(S_1^{(1)} - S_0^{(1)}) + (\rho - \rho^2)S_2^{(1)} + \dots$$

$$+ (\rho^{n-1} - \rho^n)S_n^{(1)} > 0, \text{ and by the same type of argument}$$

$$\text{it can be shown that } S_0^{(2)} < S_1^{(2)} < \dots < S_{\frac{n}{2}}^{(2)} > \dots$$

$> S_n^{(2)}$ by using only the elements of $R^{(n)}$ and the ordering of the $S_j^{(1)}$'s. Hence, by induction, the same ordering holds for each power r and the lemma follows by taking the limit.

Lemma 2.2: Under the hypothesis of lemma 1, $\phi_0^{(n)}$

$$= \max_j \phi_{0j}^{(n)} \rightarrow 0 \text{ as } n \rightarrow \infty.$$

Proof:

$$R^{(n)} \phi_0^{(n)} = \sigma^2 \begin{bmatrix} 1 & \rho & \rho^2 & \dots & \dots & \dots & \rho^{n-1} \\ \rho & 1 & \rho & \dots & \dots & \dots & \rho^{n-1} \\ \vdots & \vdots & \vdots & \ddots & \ddots & \ddots & \vdots \\ \rho^{\frac{n}{2}} & \rho^{\frac{n}{2}-1} & \dots & \rho 1 \rho & \dots & \rho^{\frac{n}{2}} \\ \vdots & \vdots & \vdots & \vdots & \ddots & \ddots & \vdots \\ \rho^n & \rho^{n-1} & \dots & \dots & \dots & \rho 1 \end{bmatrix} \begin{bmatrix} \phi_{00}^{(n)} \\ \phi_{01}^{(n)} \\ \vdots \\ \phi_{0\frac{n}{2}}^{(n)} \\ \vdots \\ \phi_{00}^{(n)} \end{bmatrix}$$

$$= \begin{bmatrix} a_0^{(n)} & \phi_{00}^{(n)} \\ a_1^{(n)} & \phi_{01}^{(n)} \\ \vdots & \vdots \\ a_{\frac{n}{2}}^{(n)} & \phi_{0\frac{n}{2}}^{(n)} \\ \vdots & \vdots \\ a_0^{(n)} & \phi_{00}^{(n)} \end{bmatrix} \quad \text{where}$$

$$a_0^{(n)} = \sigma^2 \left[1 + \rho \frac{\phi_{01}^{(n)}}{\phi_{00}^{(n)}} + \rho^2 \frac{\phi_{02}^{(n)}}{\phi_{00}^{(n)}} + \dots + \rho^{n-1} \frac{\phi_{01}^{(n)}}{\phi_{00}^{(n)}} + \rho^n \right]$$

$$a_{\frac{n}{2}}^{(n)} = \sigma^2 \left[1 + 2\rho \frac{\phi_{0 \frac{n}{2}-1}^{(n)}}{\phi_{0 \frac{n}{2}}^{(n)}} + 2\rho^2 \frac{\phi_{0 \frac{n}{2}-2}^{(n)}}{\phi_{0 \frac{n}{2}}^{(n)}} + \dots + 2\rho^{\frac{n}{2}-1} \frac{\phi_{01}^{(n)}}{\phi_{0 \frac{n}{2}}^{(n)}} + 2\rho^{\frac{n}{2}} \frac{\phi_{00}^{(n)}}{\phi_{0 \frac{n}{2}}^{(n)}} \right].$$

But since $\phi_0^{(n)}$ is an eigenvector $a_j^{(n)} = \lambda_0^{(n)}$,
 $j = 0, \dots, \frac{n}{2}$. Let $\tilde{\lambda}_0^{(n)} = \frac{\lambda_0^{(n)}}{\sigma^2}$, then since $\lambda_0^{(n)} \rightarrow$

$$r_x(0) = \frac{1}{(1-\rho)^2}, \quad \lambda_0^{(n)} \rightarrow \frac{1}{\sigma^2(1-\rho)^2} = \frac{(1-\rho^2)}{(1-\rho)^2}$$

$$= \frac{(1-\rho)(1+\rho)}{(1-\rho)^2} = \frac{1+\rho}{1-\rho} = 1 + 2\rho + 2\rho^2 + \dots. \text{ In the}$$

expression for $a_{\frac{n}{2}}^{(n)}$ it follows from lemma 1 that the

coefficient of each positive power of ρ is ≤ 2 . Hence

in order for $a_{\frac{n}{2}}^{(n)}$ to approach $1 + 2\rho + 2\rho^2 + \dots$ it

follows that for each fixed k , $\phi_{\frac{n}{2}-k}^{(n)} \rightarrow 1$. On the

other hand since $\sum_{v=0}^n \phi_{0v}^{(n)2} = 1$, then $\phi_{\frac{n}{2}}^{(n)} \rightarrow 0$ as

$n \rightarrow \infty$.

Theorem 2.1: Let $\phi_0^{(n)}$ be as in the preceding lemmas,

and let $g_0^{(n)}(\omega) = \sum_{v=0}^n \phi_{0v}^{(n)} e^{-iv\omega}$. Then $|g_0^{(n)}(\omega)|^2$

$\rightarrow \delta(\omega)$ as $n \rightarrow \infty$.

Proof: $|g_0^{(n)}(\omega)|^2 = \sum_{k=-n}^n c_k e^{ik\omega}$, where $c_k = c_{-k}$

$= \sum_{v=k}^n \phi_{0v}^{(n)} \phi_{0v-k}^{(n)}$ so that in particular $c_0 = \sum_{v=0}^n \phi_{0v}^{(n)2}$

$$\begin{aligned}
&= 1. \text{ For } k < \frac{n}{2}, \sum_{j=0}^n \phi_{0j}^{(n)^2} - \sum_{j=k}^n \phi_{0j}^{(n)} \phi_{0j-k}^{(n)} \\
&= \phi_{00}^{(n)^2} + \dots + \phi_{0k-1}^{(n)^2} + \phi_{0k}^{(n)} (\phi_{0k}^{(n)} - \phi_{00}^{(n)}) \\
&\quad + \phi_{0k+1}^{(n)} (\phi_{0k+1}^{(n)} - \phi_{01}^{(n)}) + \dots + \phi_{0 \frac{n}{2}}^{(n)} (\phi_{0 \frac{n}{2}}^{(n)} - \phi_{0 \frac{n}{2}-k}^{(n)}) \\
&\quad + \phi_{0 \frac{n}{2}+1}^{(n)} (\phi_{0 \frac{n}{2}+1}^{(n)} - \phi_{0 \frac{n}{2}+1-k}^{(n)}) + \dots + \phi_{0 n-1}^{(n)} (\phi_{0 n-1}^{(n)} - \phi_{0 n-k-1}^{(n)}) \\
&\quad + \phi_{0n}^{(n)} (\phi_{0n}^{(n)} - \phi_{0n-k}^{(n)}). \text{ But } \phi_{00} = \phi_{0n}, \phi_{01} = \phi_{0 n-1}, \\
&\quad \dots, \phi_{0 \frac{n}{2}-1} = \phi_{0 \frac{n}{2}+1}, \text{ implies } \phi_{0k} (\phi_{0k} - \phi_{00}) \\
&\quad + \phi_{0n} (\phi_{0n} - \phi_{0n-k}) = \phi_{0k} (\phi_{0k} - \phi_{00}) - \phi_{00} (\phi_{0k} - \phi_{00}) \\
&= (\phi_{0k} - \phi_{00})^2, \text{ etc., so that the above sum becomes} \\
&\quad \sum_{j=0}^n \phi_{0j}^2 - \sum_{j=k}^n \phi_{0j} \phi_{0j-k} = \sum_{j=0}^{k-1} \phi_{0j}^2 + \sum_{j=k}^{n/2} (\phi_{0j} - \phi_{0j-k})^2 \\
&= S_1 + S_2 > 0. \text{ Since } \phi_{0 \frac{n}{2}}^{(n)} \rightarrow 0, \text{ then for fixed } k, \\
&\quad S_1 \leq k (\phi_{0 \frac{n}{2}}^{(n)})^2 \rightarrow 0 \text{ as } n \rightarrow \infty. \text{ Let } r \text{ be the greatest in-}
\end{aligned}$$

teger such that $rk \leq \frac{n}{2}$, and let $a_1 = \min(rk+1, \frac{n}{2})$, a_2
 $= \min(rk+2, \frac{n}{2}), \dots, a_{k-1} = \min((r+1)k-1, \frac{n}{2})$. Then S_2
 $= \sum_{j=k}^{n/2} (\phi_{0j} - \phi_{0j-k})^2 \leq (\phi_{0k} - \phi_{00})^2 + (\phi_{02k} - \phi_{0k})^2 + \dots$
 $+ (\phi_{0rk} - \phi_{0(r-1)k})^2 + (\phi_{0 \frac{n}{2}} - \phi_{0rk})^2 + (\phi_{0k+1} - \phi_{01})^2$
 $+ \dots + (\phi_{0a_1} - \phi_{0(r-1)k+1})^2 + (\phi_{0 \frac{n}{2}} - \phi_{0a_1})^2. \dots$
 $+ (\phi_{02k-1} - \phi_{0k-1})^2 + \dots + (\phi_{0a_{k-1}} - \phi_{0rk-1})^2$
 $+ (\phi_{0 \frac{n}{2}} - \phi_{0a_{k-1}})^2$. Now, each term of $(\phi_{0k} - \phi_{00})$
 $+ (\phi_{02k} - \phi_{0k}) + \dots + (\phi_{0rk} - \phi_{0(r-1)k}) + (\phi_{0 \frac{n}{2}} - \phi_{0rk})$
 $= (\phi_{0 \frac{n}{2}} - \phi_{00})$ is non-negative so that the sum of squares
 of these terms is less than or equal to the square of the
 sum. The same is true for each of the other $k-1$ sums so
 that $S_2 \leq k(\phi_{0 \frac{n}{2}} - \phi_{00})^2 \rightarrow 0$ as $n \rightarrow \infty$. But this implies
 that for each fixed k , $c_k = c_{-k} + 1$ as $n \rightarrow \infty$. From this

it follows that for any positive integer N , and any $\epsilon > 0$, there exists an integer M such that $n > M$

$$\Rightarrow |1 - c_k^{(n)}| < \frac{\epsilon}{2N+1} \text{ for } k = 0, 1, \dots, N. \text{ Then}$$

$$\left| \sum_{k=-N}^N c_k^{(n)} e^{ik\omega} - \sum_{k=-N}^N e^{ik\omega} \right| \leq \sum_{k=-N}^N |c_k^{(n)} - 1| < \epsilon.$$

Thus, if the partial sums $S_N(\omega) = \sum_{k=-N}^N e^{ik\omega}$ of the

trigonometrical series $\sum_{k=-\infty}^{\infty} e^{ik\omega}$ converge to a general-

ized function $g(\omega)$, the transfer function $|g_0^{(n)}(\omega)|^2$

$= \sum_{k=-n}^n c_k^{(n)} e^{ik\omega}$ must converge to the same function.

It is shown, however, in Lighthill [6, p. 67] that the

trigonometrical series $\sum_{k=-\infty}^{\infty} e^{ik\omega}$ converges to the gen-

eralized function $g(\omega) = 2\pi \sum_{k=-\infty}^{\infty} \delta(\omega - 2k\pi)$, (a train of

delta functions). When ω is restricted to the interval $[-\pi, \pi]$, $g(\omega)$ is simply $2\pi\delta(\omega)$, and it then follows that

$$|g_0^{(n)}(\omega)|^2 \xrightarrow{n \rightarrow \infty} 2\pi\delta(\omega).$$

Now suppose $\rho < 0$. Let the covariance matrices and spectral density functions now be denoted by $R^{(n)-}$ and $f_x^-(\omega)$. Since $f_x^-(\omega) = f_x(\omega + \pi)$, it follows from the theory of Toeplitz forms that the eigenvalues of $R^{(n)-}$ are the same as those of $R^{(n)}$.

Theorem 2.2: Let $g_0^{(n)-}(\omega) = \sum_{v=0}^n \phi_{0v}^{(n)-} e^{-iv\omega}$, where

$\phi_{0v}^{(n)-}$ is the v th element of $\phi_0^{(n)-}$, the eigenvector corresponding to the largest eigenvalue $\lambda_0^{(n)}$ of $R^{(n)-}$.

Then $g_0^{(n)-}(\omega) \rightarrow 0$ as $n \rightarrow \infty$ for any fixed ω such that $|\omega| \neq \pi$.

Proof: By direct multiplication, it may be verified

that $\phi_{0v}^{(n)-} = (-1)^v \phi_{0v}^{(n)}$ so that

$$g_0^{(n)-}(\omega) = \sum_{v=0}^n (-1)^v \phi_{0v}^{(n)} e^{-iv\omega} = \sum_{v=0}^n \phi_{0v}^{(n)} e^{-i(\omega \pm \pi)v} \rightarrow 0$$

whenever $(\omega \pm \pi) \neq 0$.

These same results can be obtained to show the behavior of the first principal component process of certain higher order moving average and autoregressive processes. For instance, consider the second order moving average process $x(t) = Z(t) + \alpha_1 Z(t-1) + \alpha_2 Z(t-2)$, where $Z(t)$ is a white noise process, $0 < \alpha_1 < 1$, $0 < \alpha_2 < 1$. $x(t)$ has spectral density function $f_x(\omega) = \frac{1}{2\pi} (1 + \alpha_1^2 + \alpha_2^2 + 2(\alpha_1 + \alpha_1\alpha_2)\cos\omega + 2\alpha_2\cos2\omega)$ which attains its maximum at $\omega = 0$. The covariance matrix in this case is

$$R(n) = \begin{bmatrix} 1+\alpha_1^2+\alpha_2^2 & \alpha_1+\alpha_1\alpha_2 & \alpha_2 & 0 & \dots & 0 \\ \alpha_1+\alpha_1\alpha_2 & 1+\alpha_1^2+\alpha_2^2 & \alpha_1+\alpha_1\alpha_2 & \alpha_2 & 0 & \dots & 0 \\ \alpha_2 & \alpha_1+\alpha_1\alpha_2 & 1+\alpha_1^2+\alpha_2^2 & \alpha_1+\alpha_1\alpha_2 & \alpha_2 & 0 & \dots & 0 \\ 0 & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots & \vdots \\ 0 & \dots & \dots & \dots & \alpha_2 & \alpha_1+\alpha_1\alpha_2 & 1+\alpha_1^2+\alpha_2^2 \end{bmatrix}$$

The results of Perron's theorem and its corollary can

$$= \begin{pmatrix} a_0^{(n)} & \phi_0 \\ a_1^{(n)} & \phi_1 \\ \vdots & \vdots \\ a_{\frac{n}{2}-2}^{(n)} & \phi_{\frac{n}{2}-2} \\ a_{\frac{n}{2}-1}^{(n)} & \phi_{\frac{n}{2}-1} \\ a_{\frac{n}{2}}^{(n)} & \phi_{\frac{n}{2}} \\ \vdots & \vdots \end{pmatrix},$$

$$\text{where } a_0^{(n)} = \alpha + \beta \frac{\phi_1}{\phi_0} + \gamma \frac{\phi_2}{\phi_0}$$

$$a_1^{(n)} = \alpha + \beta \left(\frac{\phi_0}{\phi_1} + \frac{\phi_2}{\phi_1} \right) + \gamma \left(\frac{\phi_2}{\phi_1} \right)$$

$$a_{\frac{n}{2}-2}^{(n)} = \alpha + \beta \left[\frac{\phi_{\frac{n}{2}-1}}{\phi_{\frac{n}{2}-2}} + \frac{\phi_{\frac{n}{2}-3}}{\phi_{\frac{n}{2}-2}} \right] + \gamma \left[\frac{\phi_{\frac{n}{2}}}{\phi_{\frac{n}{2}-2}} + \frac{\phi_{\frac{n}{2}-4}}{\phi_{\frac{n}{2}-2}} \right]$$

$$a_{\frac{n}{2}-1}^{(n)} = \alpha + \beta \left[\frac{\phi_{\frac{n}{2}}}{\phi_{\frac{n}{2}-1}} + \frac{\phi_{\frac{n}{2}-2}}{\phi_{\frac{n}{2}-1}} \right] + \gamma \left[1 + \frac{\phi_{\frac{n}{2}-3}}{\phi_{\frac{n}{2}-1}} \right]$$

$$a_{\frac{n}{2}}^{(n)} = \alpha + 2\beta \frac{\phi_{\frac{n}{2}-1}}{\phi_{\frac{n}{2}}} + 2\gamma \frac{\phi_{\frac{n}{2}-2}}{\phi_{\frac{n}{2}}},$$

and $\alpha = 1 + \alpha_1^2 + \alpha_2^2$, $\beta = \alpha_1 + \alpha_1\alpha_2$, $\gamma = \alpha_1$. Since

ϕ_0 is an eigenvector, $a_0^{(n)} = a_1^{(n)} = \dots = a_{\frac{n}{2}}^{(n)}$

$= \lambda_0^{(n)}$ and as $n \rightarrow \infty$, $\lambda_0^{(n)} \rightarrow f_x(0) = \alpha + 2\beta + 2\gamma$.

The expression for $a_{\frac{n}{2}}^{(n)}$ thus shows that as $n \rightarrow \infty$,

$\phi_{\frac{n}{2}-1} / \phi_{\frac{n}{2}} \rightarrow 1$ and $\phi_{\frac{n}{2}-2} / \phi_{\frac{n}{2}} \rightarrow 1$. Working backwards then,

it follows that for any fixed k , $\phi_{\frac{n}{2}-k} / \phi_{\frac{n}{2}} \rightarrow 1$ as $n \rightarrow \infty$

and the lemma follows. Theorem 2.1 then holds exactly as before. If $\alpha_1 < 0$, $\alpha_2 > 0$, then $\alpha_1 + \alpha_1\alpha_2 < 0$ and the spectral density function now denoted by $f_x^-(\omega)$ assumes its maximum at $\lambda = \pm\pi$. $f_x^-(\omega) = f_x(\omega + \pi)$ so

that the eigenvalues of $R^{(n)-}$ are the same as those of $R^{(n)}$ except with different ordering, where $R^{(n)-}$ is the covariance matrix in the case with $\alpha_1 < 0$. $R^{(n)-}$

can be derived from $R^{(n)}$ by changing the algebraic sign of every other element and, hence, direct multiplication

will show that $\phi_0^{(n)-} = \begin{bmatrix} \phi_{00}^{(n)-} \\ \vdots \\ \phi_{0n}^{(n)-} \end{bmatrix}$, where $\phi_{0v}^{(n)-}$

$= (-1)^v \phi_{0v}^{(n)}$, so that theorem 2.2 holds.

Mixed Autoregressive Moving Average Process

Let $x(t)$ be a stationary process of the form

$$x(t) - \rho x(t-1) = Z(t) + \alpha Z(t-1). \quad (2.13)$$

This can be formally written as,

$$(1 - \rho B)x(t) = (1 + \alpha B)Z(t)$$

or,

$$x(t) = \frac{(1 + \alpha B)}{(1 - \rho B)} Z(t), \text{ where } B \text{ denotes } \quad (2.14)$$

the backwards lag operator. From (2.14) if $|\rho| < 1$, it follows that $x(t)$ can be expressed as the infinite moving average process

$$x(t) = Z(t) + (\rho + \alpha)Z(t-1) + \rho(\rho + \alpha)Z(t-2) + \dots$$

from which it follows that

$$r(0) = \frac{1 + 2\rho\alpha + \alpha^2}{1 - \rho^2} = \frac{\sigma}{1 - \rho^2},$$

$$r(n) = \frac{\rho^{n-1}[\alpha + \rho(\alpha^2 + \rho\alpha + 1)]}{1 - \rho^2} = \frac{\rho^{n-1}\gamma_2}{1 - \rho^2},$$

so that,

$$R^{(n)} = \begin{bmatrix} \delta & \gamma & \rho\gamma & \rho^2\gamma & \dots & \rho^{n-1}\gamma \\ \gamma & \delta & \gamma & \rho\gamma & \dots & \rho^{n-2}\gamma \\ \vdots & & & & & \\ \rho^{n-1}\gamma & \dots & \dots & \dots & \gamma & \delta \end{bmatrix} \frac{1}{1-\rho^2}$$

If $\rho, \alpha > 0$, $R^{(n)}$ is a positive matrix and if

$S_0^{(1)}, S_1^{(1)}, \dots, S_n^{(1)}$ denote its row sums, then

$$\begin{aligned} S_0^{(1)} &< S_1^{(1)} < \dots < S_{\frac{n}{2}}^{(1)} > \dots > S_n^{(1)} = S_0^{(1)}. \quad S_0^{(2)} \\ &= \delta S_0^{(1)} + \gamma S_1^{(1)} + \rho\gamma S_2^{(1)} + \dots + \rho^{\frac{n}{2}-1}\gamma S_n^{(1)} + \dots \\ &+ \rho^{n-1}\delta S_n^{(1)}, \quad S_1^{(2)} = \gamma S_0^{(1)} + \delta S_1^{(1)} + \gamma S_2^{(1)} + \dots \\ &+ \rho^{\frac{n}{2}-1}\gamma S_n^{(1)} + \dots + \rho^{n-2}\delta S_n^{(1)}. \quad \gamma < \delta \text{ since it is al-} \end{aligned}$$

ways true that $|r(t)| < r(0)$ and γ is real. Hence, it

follows that $S_1^{(2)} > S_0^{(2)}$, and in a similar way $S_{\frac{n}{2}}^{(2)}$

$> S_{\frac{n}{2}-1}^{(2)} > \dots > S_1^{(2)} > S_0^{(2)}$, etc., which implies that

lemma 2.1 holds. The spectral density function of $x(t)$

$$f_x(\omega) = \frac{1}{2\pi} \left| \frac{1 + \alpha e^{-i\omega}}{1 - \rho e^{-i\omega}} \right|^2 = \frac{1}{2\pi} \frac{1 + \alpha^2 + 2\alpha \cos \omega}{1 + \rho^2 - 2\rho \cos \omega}.$$

For fixed n , the row sum $S_{\frac{n}{2}}^{(1)}$

$$\begin{aligned}
 &= (\delta + 2\gamma + 2\rho\gamma + \dots + 2\rho^{\frac{n}{2}-1}\gamma) \frac{1}{1-\rho^2} \\
 &= \frac{1}{1-\rho^2} [1 + 2\rho\alpha + \alpha^2 + 2(\alpha + \rho(\alpha^2 + \rho\alpha + 1)) \\
 &\quad + \rho(\alpha + \rho(\alpha^2 + \rho\alpha + 1)) + \dots + 2\rho^{\frac{n}{2}-1}(\alpha + \rho(\alpha^2 + \rho\alpha + 1))] \\
 &= \frac{1}{1-\rho^2} [(1+\alpha)^2 + 2\rho(1+\alpha)^2 + \dots + 2\rho^{\frac{n}{2}-1}(1+\alpha)^2 \\
 &\quad + 2\rho^{\frac{n}{2}}(\alpha^2 + \alpha + 1) + 2\rho^{\frac{n}{2}+1}\alpha]. \text{ As } n \rightarrow \infty \text{ this becomes}
 \end{aligned}$$

$$\begin{aligned}
 &\frac{1}{1-\rho^2} (1+\alpha)^2 \left(\frac{2}{1-\rho} - 1 \right) = \frac{1}{1-\rho^2} (1+\alpha)^2 \frac{(1+\rho)}{(1-\rho)} \\
 &= \frac{(1+\alpha)^2}{(1-\rho)^2} = 2\pi f_x(0). \text{ The same argument as in the auto-} \\
 &\text{regressive case then shows that lemma 2.2 and hence} \\
 &\text{theorem 2.1 also hold in this case.}
 \end{aligned}$$

If both $\rho < 0$ and $\alpha < 0$ then $\delta > 0$, while $\gamma < 0$; $f_x^{(n)}(\omega) = f_x(\pi + \omega)$ has its maximum at $\omega = \pm\pi$; $\phi_0^{(n)-}$

$$= \begin{bmatrix} \phi^{(n)-} \\ 00 \\ \vdots \\ \phi_0^{(n)-} \end{bmatrix}, \text{ where } \phi_{0v}^{(n)-} = (-1)^v \phi_{0v}^{(n)} \text{ and hence}$$

theorem 2.2 holds.

CHAPTER III

PROCESSES WITH NONCONSTANT MEANS

Let $y(t)$ be a real valued second order stationary process with mean zero, and let $x(t) = m(t) + y(t)$, where $m(t)$ is a deterministic but unknown function. In this chapter the covariances considered will be expected sample covariances where the sample covariances are computed as if $m(t)$ were constant. Thus, each covariance will consist of a part due to $m(t)$ and a part due to $y(t)$ as follows.

Let $x(1), x(2), \dots, x(n+k)$ be a sample from the process $x(t)$ and define

$$\bar{x}(r) = \frac{1}{n} \sum_{t=k+1}^{n+k} x(t-r) = \bar{m}(r) + \bar{y}(r), \quad r=0, \dots, k \quad (3.1)$$

The sample covariance between $x(t)$ and $x(t-\tau)$ can then be written as

$$S(\tau) = \frac{1}{n} \sum_{t=k+1}^{n+k} (x(t) - \bar{x}(0))(x(t-\tau) - \bar{x}(\tau))$$

$$\begin{aligned}
&= \frac{1}{n} \sum_{t=k+1}^{n+k} (m(t) + y(t) - \bar{m}(0) - \bar{y}(0))(m(t - \tau) \\
&\quad + y(t - \tau) - \bar{m}(\tau) - \bar{y}(\tau)) \\
&= \frac{1}{n} \sum_{t=k+1}^{n+k} m(t)m(t - \tau) - \bar{m}(0)\bar{m}(\tau) \\
&\quad + \frac{1}{n} \sum_{t=k+1}^{n+k} m(t)[y(t - \tau) - \bar{y}(\tau)] \\
&\quad + \frac{1}{n} \sum_{t=k+1}^{n+k} m(t - \tau)[y(t) - \bar{y}(0)] \\
&\quad + \frac{1}{n} \sum_{t=k+1}^{n+k} y(t)y(t - \tau) - \bar{y}(0)\bar{y}(\tau) .
\end{aligned}$$

Since $y(t)$ is stationary then (approximately) $ES(\tau)$

$$= \frac{1}{n} \sum_{t=k+1}^{n+k} m(t)m(t - \tau) - \bar{m}(0)\bar{m}(\tau) + r_y(\tau), \quad (3.2)$$

If $y(t)$ is a white noise process, then the equation becomes

$$\begin{aligned}
ES(\tau) &= \frac{1}{n} \sum_{t=k+1}^{n+k} m(t)m(t - \tau) - \bar{m}(0)\bar{m}(\tau) + \sigma_y^2 \delta_{\tau 0} \\
&\quad (3.3)
\end{aligned}$$

$x(t)$ A Linear Trend Plus White Noise

Let $x(t) = m(t) + Z(t)$, where $Z(t)$ is a white noise process, $EZ(t) = 0$, $E(Z(t)Z(t-\tau)) = \sigma_Z^2 \delta_{0\tau}$, and $m(t) = a + bt$ where a and b are constants. Then $m(t)m(t-\tau) = (a+bt)(a+b(t-\tau)) = (a+bt)((a-b\tau)+bt) = aa_\tau + (ab+a_\tau b)t + b^2 t^2$

$$\frac{1}{n} \sum_{t=k+1}^{n+k} m(t)m(t-\tau) = aa_\tau + \frac{b(a+a_\tau)}{n} \sum_{t=k+1}^{n+k} t$$

$$+ \frac{b^2}{n} \sum_{t=k+1}^{n+k} t^2, \text{ while } \bar{m}(0)\bar{m}(\tau)$$

$$= \frac{1}{n} \sum_{t=k+1}^{n+k} (a+bt) \frac{1}{n} \sum_{t=k+1}^{n+k} (a+b(t-\tau))$$

$$= (a + \frac{b}{n} \sum_{t=k+1}^{n+k} t)(a_\tau + \frac{b}{n} \sum_{t=k+1}^{n+k} t)$$

$$= aa_\tau + \frac{b(a+a_\tau)}{n} \sum_{t=k+1}^{n+k} t + \frac{b^2}{n^2} [\sum_{t=k+1}^{n+k} t]^2$$

$$\text{By (3.3) } ES_\tau = \frac{b^2}{n} \sum_{t=k+1}^{n+k} t^2 - \frac{1}{n} [\sum_{t=k+1}^{n+k} t]^2 + \sigma_Z^2 \delta_{\tau 0}.$$

Hence, the expected covariance matrix $ER^{(n)}$ has the form

$$ER^{(n)} = \begin{bmatrix} \sigma^2 + \alpha & \alpha & \cdot & \cdot & \cdot & \cdot & \cdot & \alpha \\ \alpha & \sigma^2 + \alpha & \alpha & \cdot & \cdot & \cdot & \cdot & \alpha \\ \cdot & \alpha & \cdot & \cdot & \cdot & \cdot & \cdot & \alpha \\ \alpha & \cdot & \cdot & \cdot & \cdot & \cdot & \cdot & \sigma^2 + \alpha \end{bmatrix}, \text{ where}$$

α = the first term above. This is a positive circular matrix, and hence for a given n , $ER^{(n)}$ has eigenvalues

$$\lambda_v^{(n)} = \sigma^2 + \alpha + \left(\sum_{j=1}^n r_v^j \right) \alpha, \text{ where } r_v \text{ is a root of}$$

$x^{n+1} - 1 = 0$. When $r_v = 1$, the corresponding eigenvalue is $\lambda_0^{(n)} = \sigma^2 + (n+1)\alpha$, while if $r_v \neq 1$, $\lambda_v^{(n)} = \sigma^2$.

The eigenvector corresponding to $\lambda_0^{(n)}$ is $\phi_0^{(n)} = \begin{bmatrix} \frac{1}{\sqrt{n+1}} \\ \cdot \\ \cdot \\ \cdot \\ \cdot \\ \frac{1}{\sqrt{n+1}} \end{bmatrix}$,

whose transfer function is $|g_0^{(n)}(\omega)|^2$

$$= \left| \frac{1}{\sqrt{n+1}} \sum_{k=0}^n e^{-ik\omega} \right|^2. \text{ But, } \frac{1}{\sqrt{n+1}} \sum_{k=0}^n (e^{-i\omega})^k$$

$$= \frac{1}{\sqrt{n+1}} \left[\frac{1 - e^{-i\omega(n+1)}}{1 - e^{-i\omega}} \right] \xrightarrow{n \rightarrow \infty} 0 \text{ for any fixed } \omega \neq 0.$$

On the other hand since $m(t)$ is non-periodic (a linear trend in this case) it can be considered as having the Fourier representation

$$m(t) = \int_{-\pi}^{\pi} e^{it\omega} d\mu(\omega),$$

where $d\mu(\omega) = 0$ for $\omega \neq 0$. As a result, the (sample) power spectrum of $x(t)$ consists of the spectrum of $y(t)$ plus a jump at $\omega = 0$ due to $m(t)$. The form of the above transfer function shows that the first principal component tends to remove this zero frequency or trend portion of $x(t)$

$x(t)$ A Periodic Function Plus White Noise

Consider the process $x(t) = \cos(\frac{\pi}{2}t) + Z(t)$, where $Z(t)$ is a white noise process. For convenience, it will be assumed that both n (the sample size) and $N + 1$ (the order of the covariance matrix) are always taken to be divisible by four. Then,

$$\bar{m}(0) = \frac{1}{n} \sum_{t=k+1}^{n+k} \cos(\frac{\pi}{2}t) = 0,$$

$$\frac{1}{N} \sum m(t)m(t) = \frac{1}{N} (0 + 1 + \dots) = 1/2,$$

$$\frac{1}{N} \sum m(t)m(t-1) = 0,$$

$$\frac{1}{N} \sum m(t)m(t-2) = \frac{1}{N} (0 - 1 + 0 \dots) = -1/2$$

$$\frac{1}{N} \sum m(t)m(t-3) = 0, \text{ etc.}$$

Hence, the expected covariance matrix is of the form

$$E R(n) = \begin{bmatrix} \sigma_z^2 + 1/2 & 0 & -1/2 & 0 & \dots & 1/2 & 0 & -1/2 & 0 \\ 0 & \sigma_z^2 + 1/2 & 0 & \dots & \dots & \dots & \dots & -1/2 \\ \vdots & & & & & & & \\ \vdots & & & & & & & \\ 0 & \dots & \dots & \dots & \dots & \dots & 0 & \sigma_z^2 + 1/2 \end{bmatrix}$$

This, again, is a circular matrix with eigenvalues

$$\lambda_v = (\sigma_z^2 + 1/2) + 1/2 \sum_{j=1}^{N+1} (-1)^j r_v^{2j}, \text{ where } r_v \text{ is}$$

a root of $x^{N+1} - 1 = 0$. The expression for λ_v assumes its maximum value when $(r_v)^{2j} = (-1)^j$ or when $r_v = \pm 1$.

Since r_v can be expressed in the form $r_v = e^{i \frac{2\pi v}{N+1}}$,

$v = 0, 1, \dots, N$. $\lambda(\text{max.})$ is attained for $v = \frac{N+1}{4}$ and

$v = \frac{3(N+1)}{4}$. Using the same notation as in Chapter II, two orthogonal eigenvectors corresponding to the double root $\lambda(\text{max.})$ are given by

$$\phi_0^{(N)} = \frac{1}{\sqrt{N+1}} \begin{bmatrix} 1 \\ e^{i\frac{\pi}{2}} \\ \vdots \\ e^{i\frac{\pi N}{2}} \end{bmatrix} \quad \text{and} \quad \phi_0^{(N)} = \frac{1}{\sqrt{N+1}} \begin{bmatrix} 1 \\ e^{i\frac{3\pi}{2}} \\ \vdots \\ e^{i\frac{3\pi N}{2}} \end{bmatrix}.$$

Let $\psi^{(N)}$ be an arbitrary linear combination $\psi^{(N)} = a\phi_0^{(N)} + b\phi_0^{(N)}$. The transfer function corresponding to $\psi^{(N)}$ is

$$|g^{(N)}(\omega)|^2 = \left| \sum_{k=0}^N \left(\frac{a}{\sqrt{N+1}} e^{i\frac{\pi}{2}k} + \frac{b}{\sqrt{N+1}} e^{i\frac{3\pi}{2}k} \right) e^{-ik\omega} \right|^2,$$

$$\text{where } g^{(N)}(\omega) = \frac{a}{\sqrt{N+1}} \sum_{k=0}^N e^{i(\frac{\pi}{2}-\omega)k} + \frac{b}{\sqrt{N+1}} \sum_{k=0}^N e^{i(\frac{3\pi}{2}-\omega)k} \rightarrow 0$$

with N for any value of ω different from $\pi/2$ or $3\pi/2$.

For $t = 0, 1, \dots$, $\cos(\frac{\pi}{2})t = \cos(\frac{3\pi}{2})t$, so that any principal component formed from an eigenvector corresponding to $\lambda(\max.)$ will tend to filter out the deterministic function $m(t)$.

APPENDIX

Figures 1 and 2 of chapter II showed the transfer functions corresponding to the two eigenvectors of the 2×2 , ($n = 1$) covariance matrix of the moving average process $x(t) = Z(t) + .9Z(t - 1)$. Figures 3 and 4 show how these transfer functions change as n increases. Figure 3 shows the transfer functions corresponding to the eigenvectors ϕ_0, ϕ_2, ϕ_4 when $n = 4$, while figure 4 shows the transfer functions corresponding to ϕ_0, ϕ_3, ϕ_6 for $n = 6$.

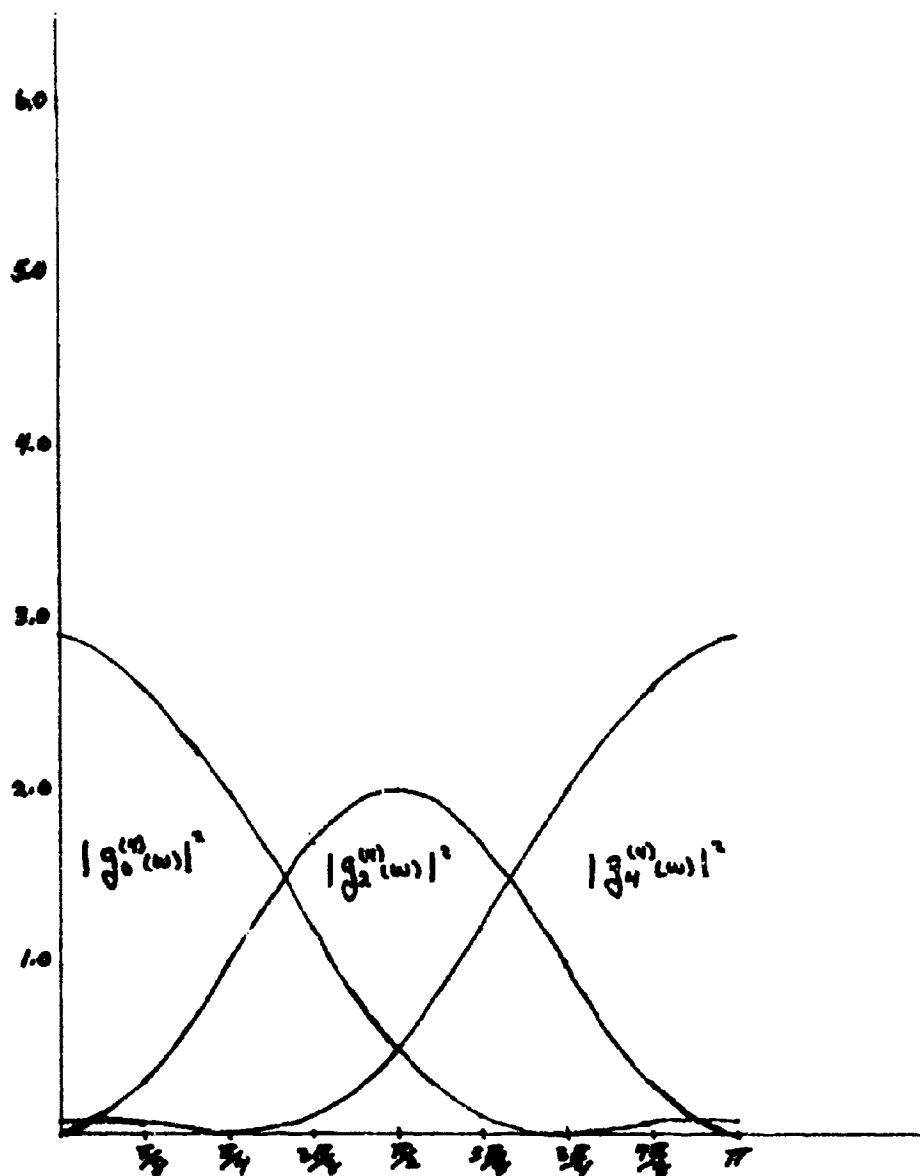


Figure 3

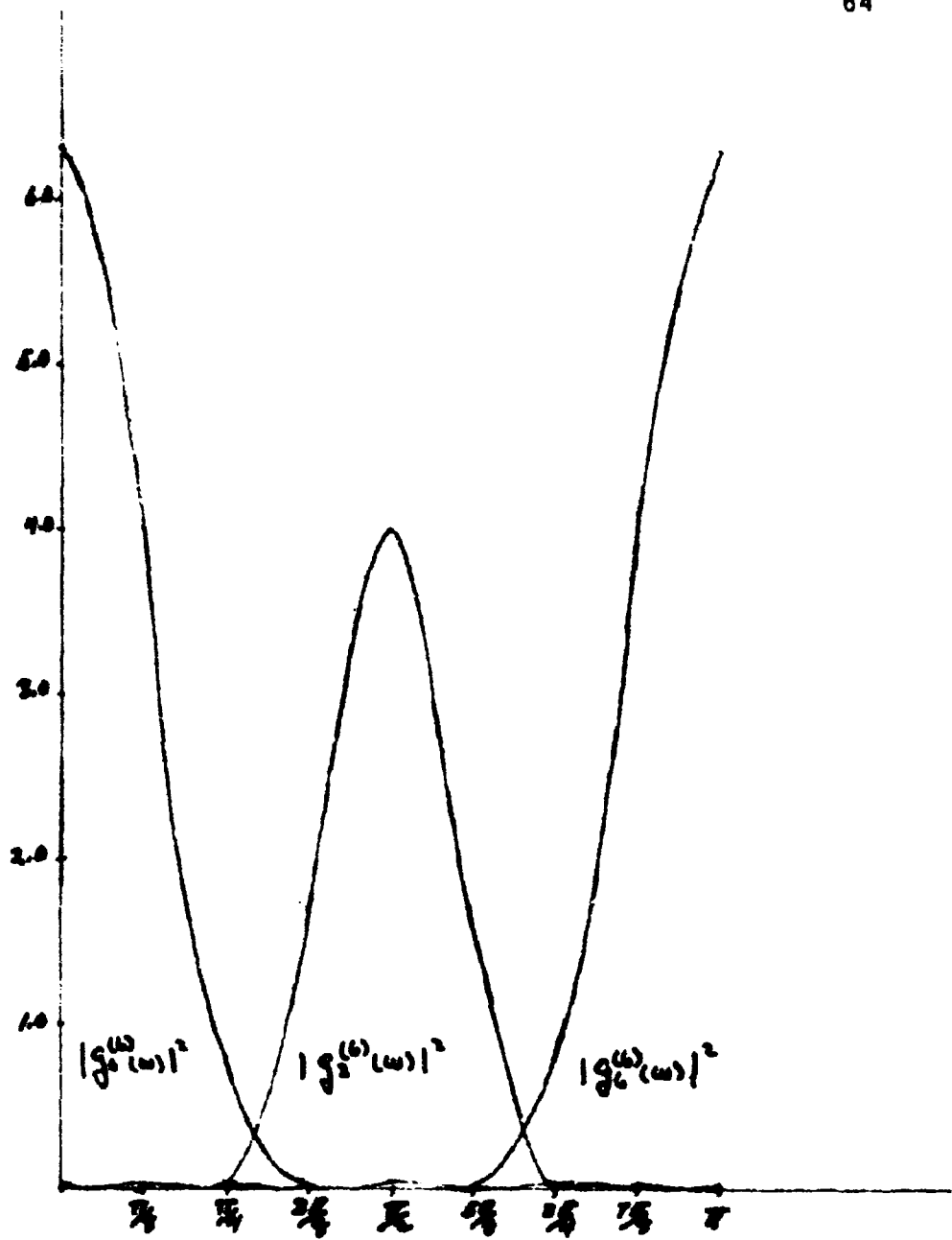


Figure 4

Figure 5 shows an example of the behavior of an autoregressive process $x(t) + .9x(t - 1) = Z(t)$ together with its first and fifth principal component processes for the case $n = 4$. The series $Z(t)$ consists of a series of random numbers taken from the Rand table of Normal Random Deviates. From this series, the series $y(t)$ was constructed. The principal component processes were constructed by numerically determining the eigenvectors of the theoretical covariance matrix of $y(t)$ and applying the appropriate filters to $y(t)$. The first principal component process can be seen to consist primarily of oscillations with period two, while the fifth principal component process contains oscillations of much lower frequencies.

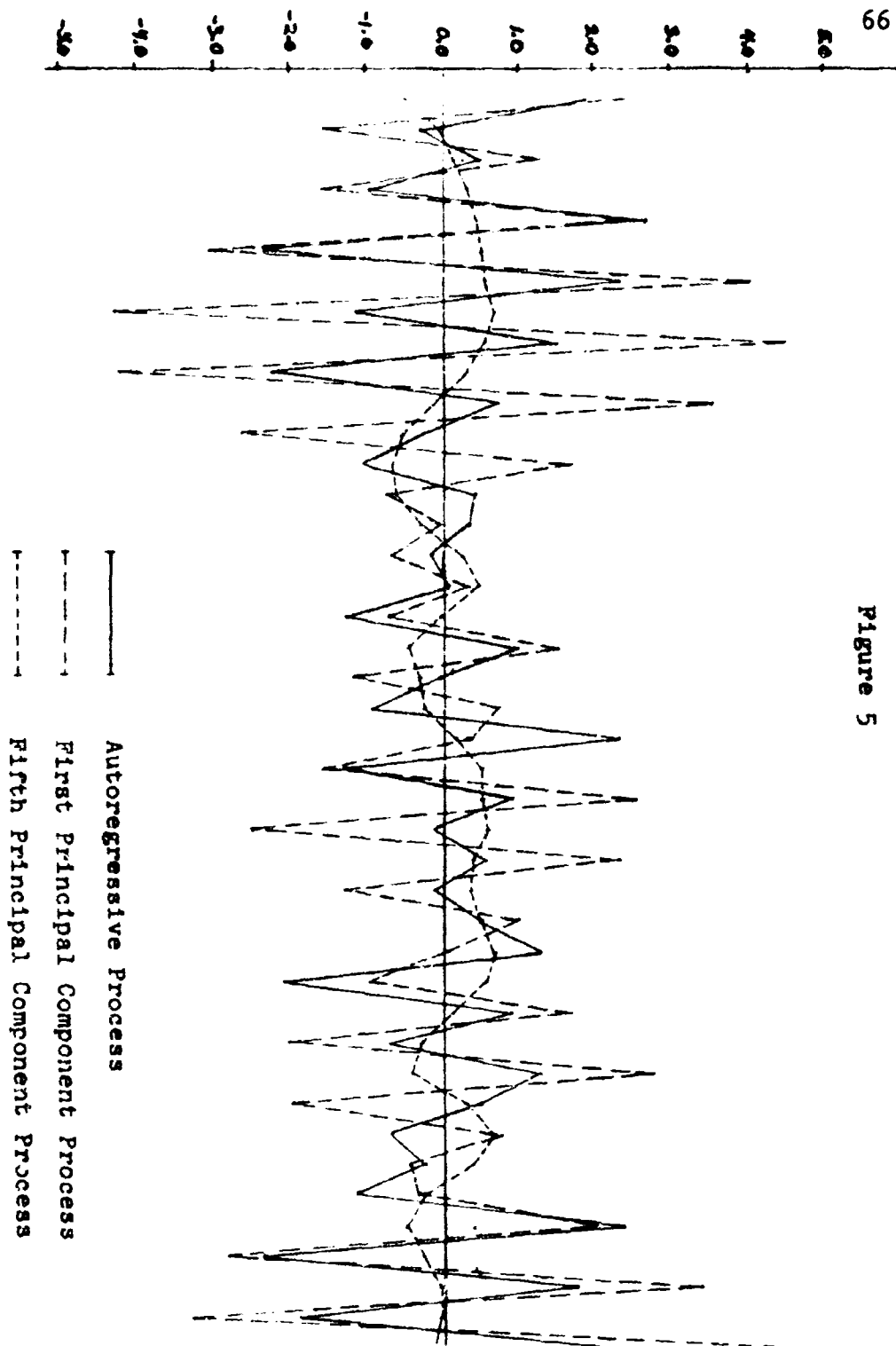


Figure 6 shows a sample of the process $x(t)$ $= \frac{1}{10}t + 5\cos\frac{\pi}{2}t + y(t)$, where $y(t) + .9y(t-1) = Z(t)$, $Z(t)$ a white noise process. Also shown are the first and second principal component processes for $n = 18$. The process $x(t)$ was constructed by taking as $Z(t)$, two hundred numbers from the Rand Table of Random Normal Deviates, constructing $y(t)$, and combining $y(t)$ with $\frac{1}{10}t$ and $5\cos\frac{\pi}{2}t$. The data was keypunched and the principal components determined (from the sample covariance matrix) using the BMDOIM principal component program. This process was selected as an example since it might resemble an unseasonally adjusted economic series.

The first principal component contains the linear trend plus some low frequency oscillations, while the second principal component seems to consist almost entirely of oscillations with period four corresponding to the cosine term. The third principal component (not shown) also consists primarily of oscillations with period four as would be expected from the results of chapter III, while the fourth (also not shown) component contains oscillations of period two from the autoregressive process.

Figure 6

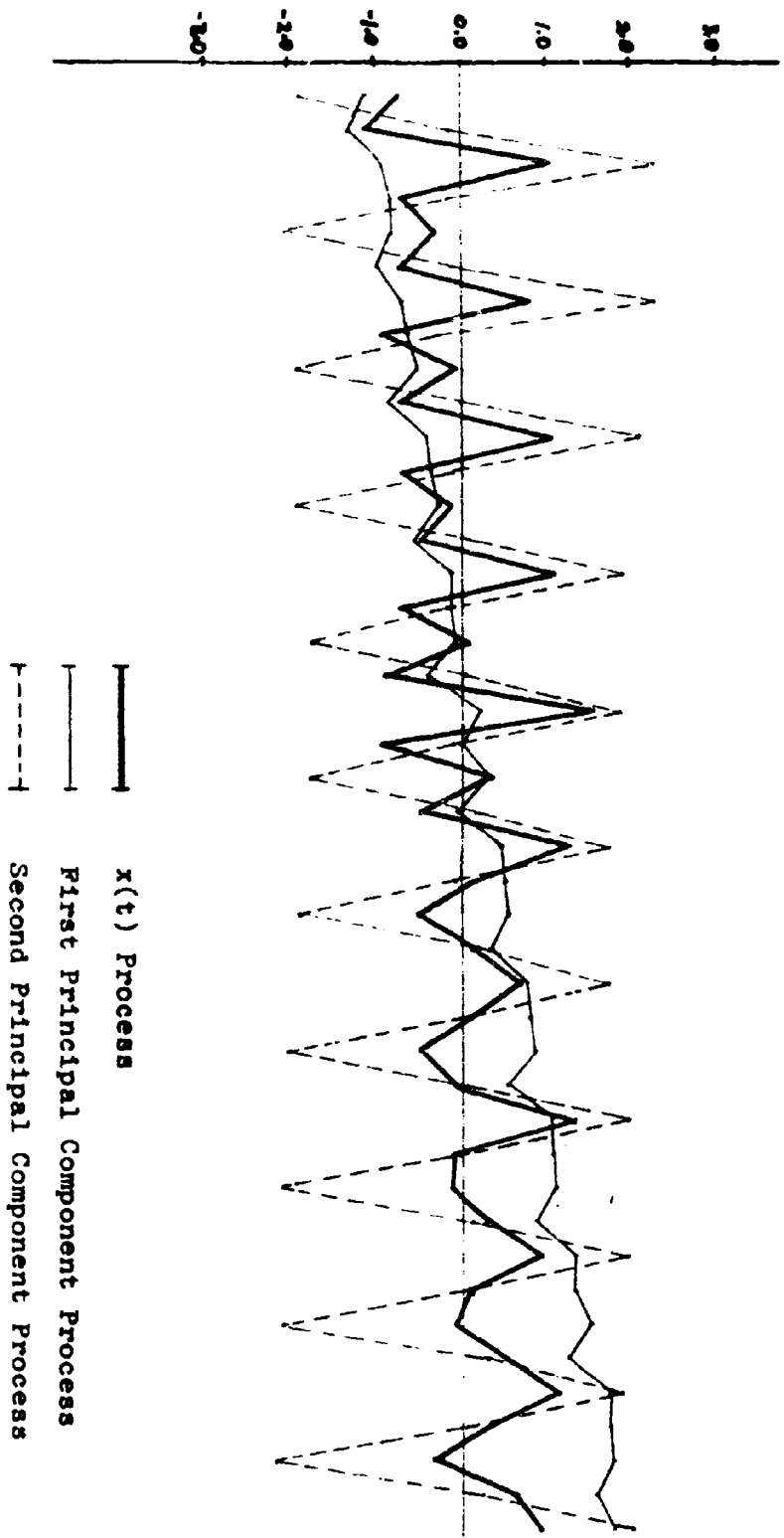
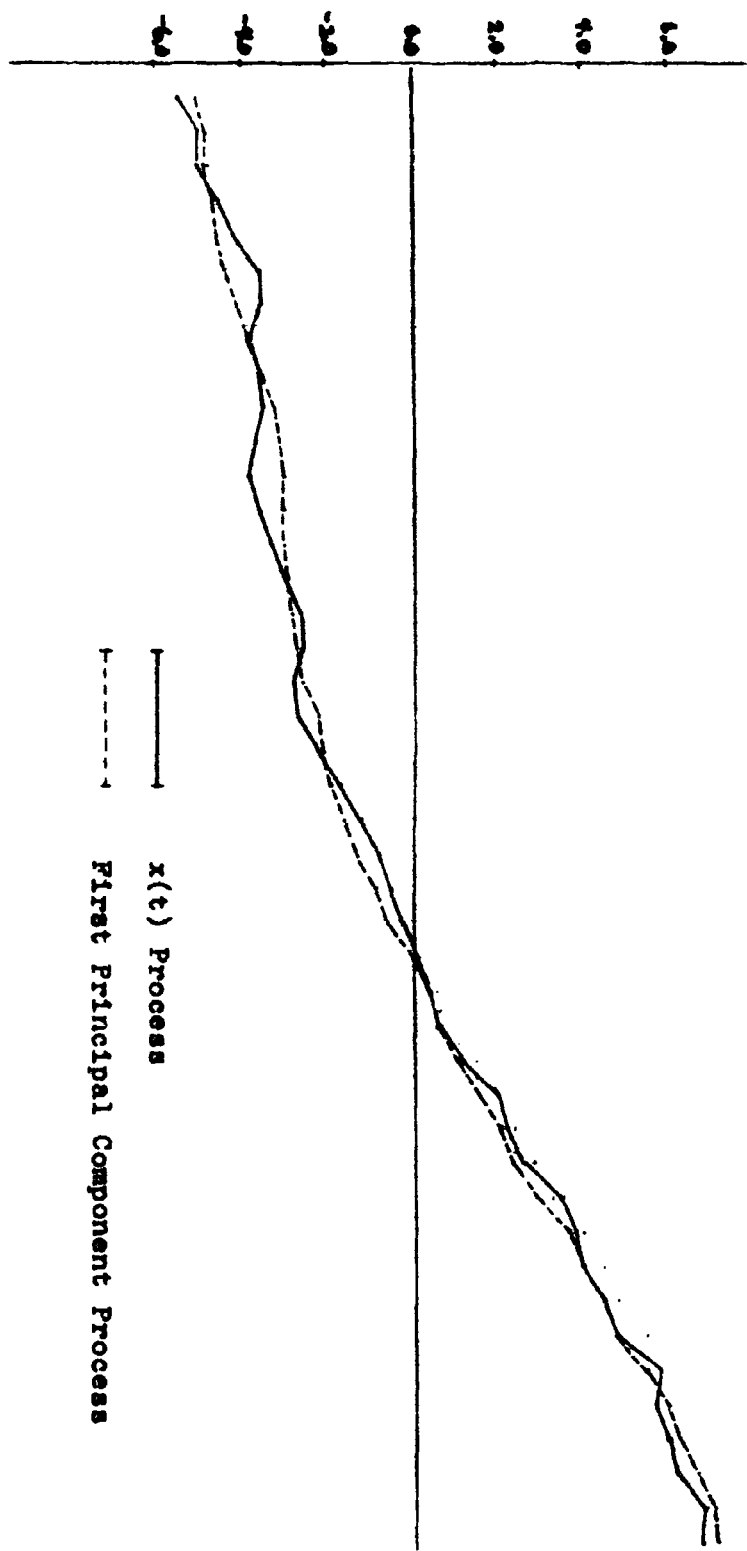
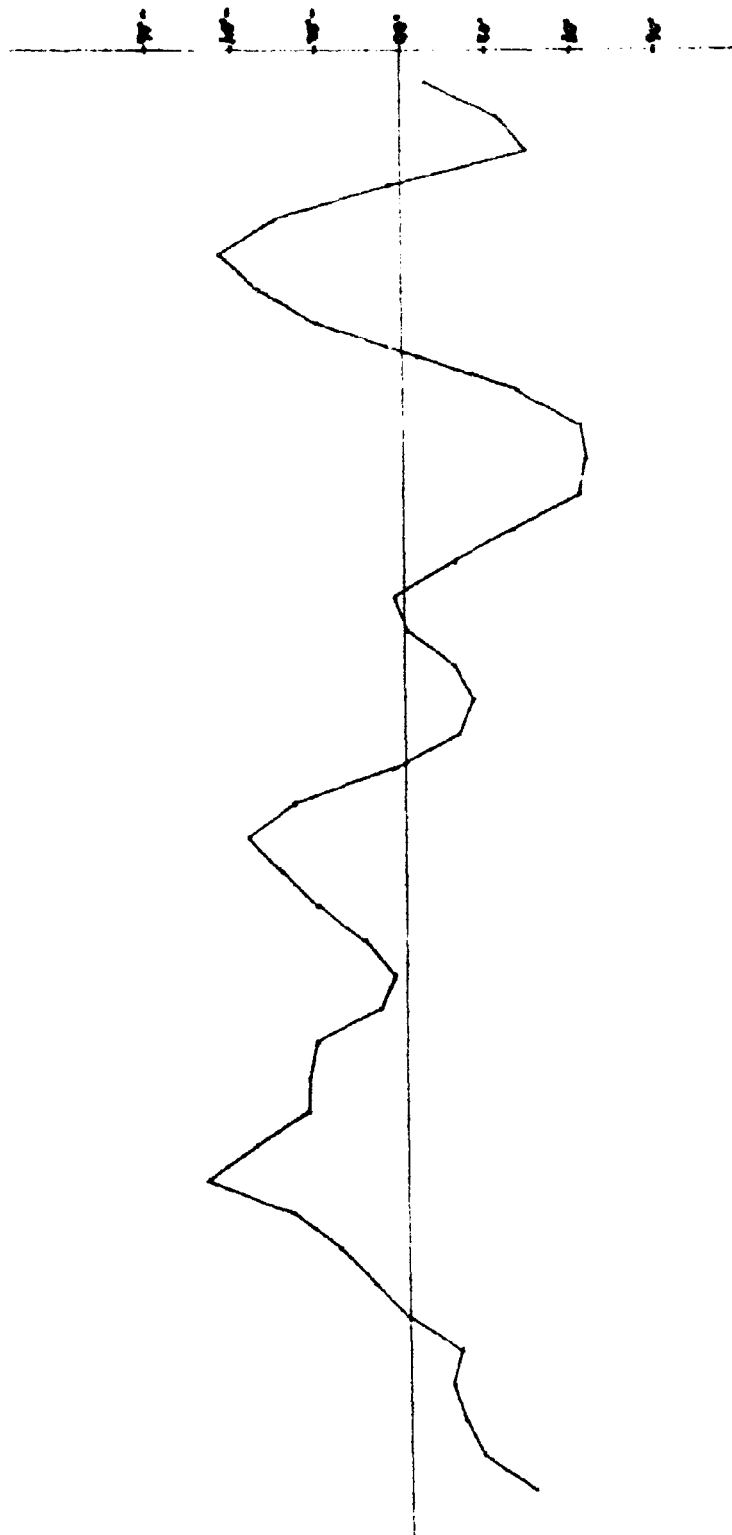


Figure 7 shows a series $x(t) = \log_e$ (the total U.S. Personal Income) , a quarterly, seasonally adjusted series from "Survey of Current Business", from first quarter 1958 through fourth quarter 1968. Shown with $x(t)$ is its first principal component process multiplied by a scale factor so that it fits the original series. These principal components were also computed using the BMD01M program. In this example, n was chosen to be equal to six. Figure 8 shows the second principal component on a greatly magnified scale.

Figure 7





71

Figure 8

BIBLIOGRAPHY

- [1] Anderson, T.W. An Introduction to Multivariate Statistical Analysis. New York and London. John Wiley & Sons, Inc., 1958.
- [2] Bellman, Richard.. Introduction to Matrix Analysis. New York, Toronto, and London: McGraw-Hill Book Company, Inc., 1960.
- [3] Cramér, Harold, and Leadbetter, M.R. Stationary and Related Stochastic Processes. New York, London, and Sidney: John Wiley & Sons, Inc., 1967.
- [4] Franklin, Joel N. Matrix Theory. Englewood Cliffs, New York: Prentice Hall, Inc., 1968.
- [5] Grenander, Ulf and Szegő, Gabor. Toeplitz Forms and their Applications. Berkeley and Los Angeles: University of California Press, 1958.
- [6] Lighthill, M.J. Introduction to Fourier Analysis and Generalized Functions. Cambridge: Cambridge University Press, 1964.
- [7] Loève, Michel. Probability Theory. Princeton, New Jersey, Toronto, New York, and London: D. Van Nostrand Company, Inc., 1963.
- [8] Massey, William F. "Principal Components Regression in Exploratory Research," Journal of the American Statistical Association, March 1965. 234-256.
- [9] Rao, C. Radhakrishna. "The Use and Interpretation of Principal Component Analysis in Applied Research," Sankhya, ser. A, vol. 26 (1964) 329-358.

- [10] Watanabe, Satoshi. Knowing and Guessing. A Quantitative Study of Inference and Information. New York, London, Sidney, and Toronto: John Wiley & Sons, Inc., 1969.

Security Classification			DEPARTMENT CONTROL DATA - R & S		
(Security classification of this report is determined by the classification of the information contained herein and the overall report is classified)					
1. ORIGINATING ACTIVITY (Corporate Name)				2a. REPORT SECURITY CLASSIFICATION	
The Ohio State University Division of Statistics Columbus, Ohio 43210				UNCLASSIFIED	
3. REPORT TITLE				2b. GROUP	
PRINCIPAL COMPONENT ANALYSIS OF TIME SERIES					
4. DESCRIPTIVE NOTES (Type of report and inclusive dates)					
Scientific Interim					
5. AUTHOR(S) (First name, middle initial, last name)					
J. Richard Stewart					
6. REPORT DATE			7a. TOTAL NO. OF PAGES		7b. NO. OF REFS
September 1970			73		10
8a. CONTRACT OR GRANT NO.			9a. ORIGINATOR'S REPORT NUMBER(S)		
AF-AFOSR 1305-67					
b. PROJECT NO.			9b. OTHER REPORT NO(S) (Any other numbers that may be assigned to this report)		
9749					
c. 61102F			AFOSR 70-2809TR		
d. 681304					
10. DISTRIBUTION STATEMENT					
1. This document has been approved for public release and sale; its distribution is unlimited.					
11. SUPPLEMENTARY NOTES			12. SPONSORING MILITARY ACTIVITY		
TECH, OTHER			Air Force Office of Scientific Research (NM) 1400 Wilson Boulevard Arlington, Virginia 22209		
13. ABSTRACT					
<p>The primary purpose of this dissertation is to investigate the properties of the principal components of a finite set of random variables comprising a part of a discrete time series. In the first chapter, the covariance structure between a set of random variables y, x_1, \dots, x_p which yields the result that the first k ($k < p$) principal components of x_1, \dots, x_p provide a better predictor of y (in the sense of expected squared error) than do any k of the variables x_1, \dots, x_p themselves, is examined.</p> <p>In the remaining chapters, principal component processes which are linear combinations of $x(t), x(t-1), \dots, x(t-n)$ where $x(t)$ is a random process and n is an arbitrary positive integer, are defined and their properties investigated in terms of their frequency content. It is shown that when $x(t)$ is a stationary moving average process, an autoregressive process, or a mixed moving average autoregressive process, the first principal component process tends (as $n \rightarrow \infty$) to contain only the frequency at which the spectral density of $x(t)$ obtains its maximum value. It is shown, moreover, that when the process $x(t)$ contains deterministic components such as a trend or a periodic component, certain of the principal component processes tend to model those deterministic components.</p>					

DEC 1970 1400